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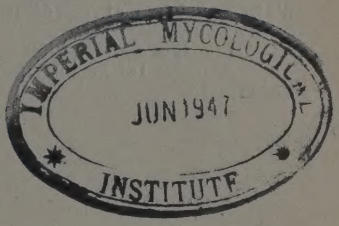

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MATHÉMATIQUES

SUR LES SÉRIES TRIGONOMÉTRIQUES UNIVERSELLES

Par D. MENCHOFF

(Présenté par A. N. Kolmogoroff, de l'Académie, le 29.VI.1945)

On peut démontrer qu'il existe une série trigonométrique

$$\frac{a_0}{2}+\sum_{n=1}^{\infty}(a_n\cos nx+b_n\sin nx) \tag{1}$$

à coefficients tendant vers zéro avec $\frac{1}{n}$, telle que chaque suite de sommes partielles

$$S_{n_k}(x)=\frac{a_0}{2}+\sum_{l=1}^{n_k}(a_l\cos lx+b_l\sin lx) \tag{2}$$

correspondant à une suite croissante quelconque de nombres entiers et positifs n_k, $k=1, 2, \ldots$, ne converge presque partout vers aucune limite finie (1).

La question se pose de savoir quelle est la classe des séries trigonométriques possédant des suites de sommes partielles $S_{n_k}(x)$, $k=1, 2, \ldots$, qui convergent presque partout. On a le théorème suivant.

Théorème 1. *Une série trigonométrique arbitraire* (1) *peut être représentée comme une somme de deux séries trigonométriques*

$$\frac{a_0'}{2}+\sum_{n=1}^{\infty}(a_n'\cos nx+b_n'\sin nx) \tag{3}$$

et

$$\frac{a_0''}{2}+\sum_{n=1}^{\infty}(a_n''\cos nx+b_n''\sin nx) \tag{4}$$

chacune desquelles possède une suite de sommes partielles convergente presque partout vers une limite finie. De plus, si

$$\lim_{n\to\infty}a_n=0,\quad \lim_{n\to\infty}b_n=0 \tag{5}$$

on peut choisir les séries (3) *et* (4) *de telle façon que*

$$\left.\begin{aligned}&\lim_{n\to\infty}a_n'=0,\quad \lim_{n\to\infty}b_n'=0\\&\lim_{n\to\infty}a_n''=0,\quad \lim_{n\to\infty}b_n''=0.\end{aligned}\right\} \tag{6}$$

Pour obtenir une généralisation du théorème 1 nous introduirons la notion d'une série trigonométrique universelle. On dit qu'une série (1) est universelle, lorsqu'il correspond à chaque fonction mesurable $f(x)$, possédant

en chaque point du segment $[0, 2\pi]$ une valeur déterminée finie ou non*, une suite de sommes partielles $S_{nk}(x)$, $k=1, 2, \ldots$ de la série considérée qui converge vers $f(x)$ presque partout dans $[0, 2\pi]$. Le théorème 1 est une conséquence du théorème suivant.

Théorème 2. *Une série trigonométrique arbitraire* (1) *peut être représentée comme une somme de deux séries trigonométriques universelles* (3) *et* (4) *dont les coefficients vérifient les conditions*

$$\left.\begin{array}{ll} \overline{\lim\limits_{n\to\infty}}\,|a'_n| \leqslant \overline{\lim\limits_{n\to\infty}}\,|a_n|, & \overline{\lim\limits_{n\to\infty}}\,|b'_n| \leqslant \overline{\lim\limits_{n\to\infty}}\,|b_n| \\ \overline{\lim\limits_{n\to\infty}}\,|a''_n| \leqslant \overline{\lim\limits_{n\to\infty}}\,|a_n|, & \overline{\lim\limits_{n\to\infty}}\,|b''_n| \leqslant \overline{\lim\limits_{n\to\infty}}\,|b_n| \end{array}\right\} \tag{7}$$

La démonstration du théorème 2 découle du lemme suivant.

Lemme. Soit $f(x)$ une fonction mesurable quelconque, finie presque partout dans $[0, 2\pi]$. On peut déterminer, pour chaque nombre positif ε et pour chaque nombre entier et positif p, un polynôme trigonométrique

$$T(x) = \sum_{n=p}^{\nu} (c_n \cos nx + d_n \sin nx) \tag{8}$$

et un ensemble mesurable E qui possèdent les propriétés suivantes;

$\alpha)$ $|f(x) - T(x)| < \varepsilon \quad (x \in E)$

$\beta)$ mes $E > 2\pi - \varepsilon, \quad E \in [0, 2\pi]$

$\gamma)$ $|c_n| < \varepsilon, \quad |d_n| < \varepsilon \quad (n = p, p+1, \ldots, \nu)$**.

En supposant que (1) soit une série trigonométrique quelconque, déterminons les séries trigonométriques (3) et (4) qui vérifient les conditions du théorème 2. Numérotons tous les polynômes trigonométriques à coefficients rationnels dans un ordre quelconque et désignons-les par

$$T_1(x), \; T_2(x), \ldots, \; T_m(x), \ldots \tag{9}$$

Déterminons par récurrence deux suites croissantes de nombres entiers non négatifs

$$\mu_0, \; \mu_1, \ldots, \; \mu_m \ldots \tag{10}$$

et

$$\nu_1, \; \nu_2, \ldots, \; \nu_m, \ldots \tag{11}$$

et les nombres réels

$$\left.\begin{array}{l} a'_0, \; a''_0, \; a'_n, \; a''_n, \; b'_n, \; b''_n \\ (n = 1, 2, \ldots) \end{array}\right\} \tag{12}$$

de la manière suivante. Posons

$$\mu_0 = 0, \; a'_0 = a_0, \; a''_0 = 0$$

et supposons que le nombre entier négatif μ_{m-1} et les nombres réels a'_n, a''_n $(n = 0, 1, 2, \ldots \mu_{m-1})$ et b'_n, b''_n $(n = 1, 2, \ldots \mu_{m-1})$*** soient déjà définis, où m est

---

* La fonction $f(x)$ peut être égale à $+\infty$ ou à $-\infty$ dans un ensemble de mesures positives

** Une démonstration de ce lemme, différente de celle de l'auteur, a été aussi donnée par A. Kolmogoroff (la démonstration n'a pas été publiée).

*** Lorsque $\mu_{m-1} = 0$, les nombres b'_n, b''_n n'existent pas.

un nombre entier et positif quelconque. Nous nous servirons des notations

$$\left.\begin{aligned} S'_n(x) &= \frac{a'_0}{2} + \sum_{l=1}^{n} (a'_l \cos lx + b'_l \sin lx) \\ S''_n(x) &= \frac{a''_0}{2} + \sum_{l=1}^{n} (a''_l \cos lx + b''_l \sin lx) \end{aligned}\right\} \tag{13}$$

lorsque les nombres a'_l, b'_l, a''_l, b''_l $(l = 1, 2, \ldots, n)$ sont déjà déterminés, et nous nous poserons de plus

$$S'_0(x) = \frac{a'_0}{2}, \quad S''_0(x) = \frac{a''_0}{2}$$

En vertu du lemme, nous pouvons définir un polynôme trigonométrique

$$\tau_m(x) = \sum_{n=\mu_{m-1}+1}^{\nu_m} (\alpha_n \cos nx + \beta_n \sin nx) \tag{14}$$

et un ensemble mesurable E_m, qui vérifient les conditions:

α') $|T_m(x) - S'_{\mu_{m-1}}(x) - \tau_m(x)| < \frac{1}{m}$ $\quad (x \in E_m)$

β') $\operatorname{mes} E_m > 2\pi - \frac{1}{m^2}$, $\quad E_m \subset [0, 2\pi]$

γ') $|\alpha_n| < \frac{1}{m^2}$, $\quad |\beta_n| < \frac{1}{m^2}$ $(n = \mu_{m-1}+1, \ldots, \nu_m)$

Posons ensuite

$$\left.\begin{aligned} &a'_n = \alpha_n, \ b'_n = \beta_n, \ a''_n = a_n - \alpha_n, \ b''_n = b_n - \beta_n \\ &(n = \mu_{m-1}+1, \ldots, \nu_m) \end{aligned}\right\} \tag{15}$$

Nous déterminons de cette façon le nombre entier et positif $\nu_m > \mu_{m-1}$ et les nombres réels a'_n, b'_n, a''_n, b''_n pour $n = \mu_{m-1}+1, \ldots, \nu_m$. En appliquant encore une fois le lemme énoncé plus haut, nous pouvons définir un polynôme trigonométrique

$$\sigma_m(x) = \sum_{n=\nu_m+1}^{\mu_m} (\alpha_n \cos nx + \beta_n \sin nx) \tag{16}$$

et un ensemble mesurable G_m, qui vérifient les conditions:

α'') $|T_m(x) - S''_{\nu_m}(x) - \sigma_m(x)| < \frac{1}{m^2}$ $\quad (x \in G_m)$

β'') $\operatorname{mes} G_m > 2\pi - \frac{1}{m^2}$, $\quad G_m \subset [0, 2\pi]$

γ'') $|\alpha_n| < \frac{1}{m^2}$, $|\beta_n| < \frac{1}{m^2}$ $\quad (n = \nu_m+1, \ldots, \mu_m)$

Posons de plus

$$\left.\begin{aligned} &a'_n = a_n - \alpha_n, \ b'_n = b_n - \beta_n, \ a''_n = \alpha_n, \ b''_n = \beta_n \\ &(n = \nu_m+1, \ldots, \mu_m) \end{aligned}\right\} \tag{17}$$

Nous définissons de cette manière le nombre entier et positif $\mu_m > \nu_m$ et les nombres réels a'_n, b'_n, a''_n, b''_n pour $n = \nu_m+1, \ldots, \mu_m$. En partant des nombres ν_0, a'_0, a''_0 nous déterminons donc de proche en proche tous les nombres entiers non négatifs (10), (11) et tous les nombres réels (12). On peut démontrer que les séries trigonométriques (3) et (4) dont les coefficients sont les nombres (12) ainsi définis vérifient toutes les conditions du théorème 2, ce qui achève sa démonstration.

Nous introduirons à présent la définition suivante. Une série trigonométrique (1) sera dite presque convergente vers une fonction $f(x)$, finie presque partout dans $[0, 2\pi]$, si l'on a les conditions:

1) Il existe une suite de sommes partielles $S_{n_k}(x)$, $k = 1, 2, \ldots$ de la série (1) qui converge vers $f(x)$ presque partout dans $[0, 2\pi]$.

2) Lorsqu'une suite de sommes partielles de la série (1) converge vers une limite finie partout dans un ensemble E de mesure positive, cette limite coïncide avec $f(x)$ presque partout dans $[0, 2\pi]$.

On peut se demander quelles sont les séries trigonométriques qui peuvent être représentées comme une somme d'un nombre fini de séries presque convergentes? La réponse est contenue en partie dans le théorème suivant.

Théorème 3. *Soient* $f_i(x)$, $i = 1, 2, 3$, *trois fonctions mesurables arbitraires, finies presque partout dans le segment* $[0, 2\pi]$, *et soit* (1) *une série trigonométrique arbitraire. On peut alors déterminer trois séries trigonométriques*

$$\frac{a_{0i}}{2} + \sum_{n=1}^{\infty} (a_{ni} \cos nx + b_{ni} \sin nx) \quad (i = 1, 2, 3)$$

dont la somme est la série (1) *et chacune desquelles est presque convergente vers la fonction correspondante* $f_i(x)$ *avec le même indice* i.

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LITTÉRATURE CITÉE

[1] D. Menchoff, C. R. Acad. Sci. URSS, XLI, No. 2, 51 (1943).

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MATHEMATICS

IRREDUCIBLE CONTINUA AND LOCAL CONNECTEDNESS

By A. M. RODNYANSKY

(Communicated by A. N. Kolmogoroff, Member of the Academy, 2.XII. 1944)

In the present paper is investigated the situation of points of local connectedness and of local non-connectedness on irreducible continua.

As is known, a space R is said to be locally connected at point x, if for every neighbourhood $U(x)$ of x there exists a neighbourhood $U_1(x)$ of the same point x, such that every point $y \in U_1(x)$ can be joined with x by a connected set lying within $U(x)$. It is easy to give an example of a plane continuum C locally connected at a point a, for which there is a neighbourhood $U(a)$ that contains no connected open in C set containing the point a. Nevertheless, the following theorems hold.

Theorem 1. *If a continuum C is irreducible between the points a and b and locally connected at x, then every neighbourhood of the point x contains a connected neighbourhood of x.*

Theorem 1'. *If a metric space C is irreducibly connected between the points a and b and locally connected at a point x, then every neighbourhood of x contains a connected neighbourhood of the point x.*

Sierpinski proved that an irreducibly connected compactum is a simple arc. The condition of compactness may, however, be weakened.

Theorem 2. *If a metric space C possessing an enumerable basis is irreducibly connected between two points and locally compact, then C is a simple arc.*

Theorem 2'. *If a metric space C with an enumerable basis is irreducibly connected between two distinct points and is locally connected, then C is a simple arc.*

In what follows we use the notations: C is a continuum irreducible between points a and b, L is the set of points of local connectedness of the continuum C, M is the complementary set $C-L$.

The structure of the components of the set M is studied in Theorems 3 and 3'.

Theorem 3. *Every component of the set M is a locally compact semi-continuum containing more than one point*. The components of the set M coincide with the semi-components of M.*

Theorem 3'. *Let K be an arbitrary component of the set M. One of the following four cases necessarily takes place:*

1. *K is a continuum.*
2. *K is a semi-continuum which can be converted into a continuum C by adding one point.*
3. *K is the sum of a continuum and a semi-continuum having non-void intersection.*
4. *K is the sum of two semi-continua having non-void intersection.*

* By a semi-continuum a metric space R is meant, any two points of which can be joined by a continuum belonging to R; the maximal semi-continuum in R containing a point x is called the semi-component of x in R.

All the semi-continua mentioned in 3 *and* 4 *can be complemented to continua by adding single points. All the four cases are essentially distinct and may really occur.*

Remark 1. The set M may be neither locally connected nor locally compact.

The following two theorems make clear the structure of the components of the set L.

Theorem 4. *Every connected subset (in particular, every component) of the set L is either a unique point of the topological image of a segment, a semi-segment or an interval (all these cases may really occur). Every point of local connectedness of the set L is a point of local compactness of the set L, but not conversely. Thereby L may be neither locally connected nor locally compact.*

The following theorem makes clear the structure of the boundary of the components of L.

Theorem 5. *Let K be an arbitrary component of the set L; H be the boundary of K.*

1. *If K is the topological image of an interval, then H is the sum of two non-overapping continua contained in distinct components of the set M.*

2. *If K is the topological image of a semi-segment, then H is either a continuum belonging to a component of the set M, or the sum of a continuum belonging to a component of M and the end of the set K (according as the end of K is an inner point, or a boundary point of K).*

As to the number of components of the sets L and K we have the following theorem.

Theorem 6. *The system of components of L having more than one point is at most enumerable, while the system of one-point components of L may be finite, enumerable or of the continuum cardinal number (all these cases may occur). The system of components of the set M which are no continua is at most enumerable, while the cardinal number of the system of all the components of M may be finite, enumerable or equal to that of continuum; all these cases may really take place.*

I am much indebted to P. S. Alexandroff for advice and to A. S. Parkhomenko for the constant attention he paid to my work.

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MATHÉMATIQUES

UN NOUVEAU CRITÈRE NÉCESSAIRE ET SUFFISANT POUR LA RÉGULARITÉ DES CHAÎNES DE MARKOFF DONT L'ENSEMBLE DES ÉTATS POSSIBLES EST CONTINU

Par T. SARYMSAKOV, de l'Académie des Sciences d'Uzbekistan

1. Considérons un certain système S d'états possibles et incompatibles de x formant un segment $[a, b]$ et qui se succèdent en suivant la loi stochastique suivante. En un certain moment initial t_0 la loi différentielle pour la probabilité d'un état $x \subset [a, b]$ est $p_0(x)$; ensuite, quel que soit le moment suivant t_h $(h = 1, 2, 3, \ldots)$ que nous considérons, la loi différentielle de la probabilité pour un système S, qui au moment t_h se trouvait dans l'état x, $x \subset [a, b]$, de passer au moment suivant t_{h+1} en un état y, $y \subset [a, b]$, est un nombre non négatif $p(x, y)$, qui ne dépend pas des états du système S aux moments $t_0, t_1, t_2, \ldots, t_{h-1}$.

Donc, l'état du système S dans l'un quelconque des moments $t_1, t_2, \ldots$ est stochastiquement complètement déterminé par son état au moment précédant et indépendamment de ces états en tous les autres moments.

La loi différentielle de la probabilité absolue de l'état x du système S au moment t_k sera désignée par $p_k(x)$. On sait bien que

$$p_k(x) = \int_a^b p_{k-1}(t)\, p(t, x)\, dt \qquad (1)$$

Au moyen de la substitution $p_k(x) = \lambda^k p(x)$, où λ est une constante, l'équation intégrale en différences (1) se réduit évidemment à l'équation intégrale linéaire homogène

$$p(x) = \mu \int_a^b p(t)\, p(t, x)\, dt \qquad (2)$$

où $\mu = 1/\lambda$.

En tenant compte de la propriété du noyau $p(t, x)$ $\int_a^b p(t, x)\, dx = 1$, on peut établir les propriétés suivantes de l'équation intégrale (2); a) $\lambda_0 = 1$ est un nombre caractéristique pour le noyau $p(t, x)$; b) tous les autres nombres caractéristiques, s'ils existent, ont un module non inférieur à 1.

Ces faits connus étant rappelés, supposons dans la suite que le noyau $p(t, x) \geqslant 0$ est un négatif et continu sur $[a, b]$.

2. Désignons par $p^{(k)}(x, y)$ la loi différentielle de la probabilité de passer de l'état x, qui correspond au moment t_h, en l'état y, qui correspond au moment t_{h+1}.

Définition. Une chaîne est nommée régulière si la probabilité $p^{(k)}(x, y)$ pour $k \to \infty$ tend uniformément sur $[a, b]$ vers une limite $p(y)$, qui ne dépend pas de x.

En 1934 M. Fréchet [1] a démontré le théorème suivant.

Pour que le cas régulier ait lieu, il faut et il suffit que les conditions suivantes soient remplies; a) que l'unité soit une racine simple de l'équation de Fredholm $D(\lambda)=0$; b) que l'unité soit le seul nombre caractéristique du noyau $p(t, x)$ dont le module est égal à 1.

En utilisant ce critère de M. Fréchet nous allons établir un nouveau critère de régularité plus commode pour les applications.

Théorème 1. *Pour la régularité d'une chaîne de Markoff il est nécessaire et suffisant qu'il existe un entier positif n_0 tel que la m-ième itération du noyau*

$$p^m(x, x) > 0 \qquad (a \leqslant x \leqslant b)$$

pour toutes les valeurs entières de m vérifiant l'inégalité $m \geqslant n_0$.

Ce théorème est une conséquence immédiate du théorème de Fréchet et du théorème suivant, qui est la généralisation finale des théorèmes de Jentzsch [2] et Gantmacher [3].

Théorème 2*. *Supposons que dans l'équation intégrale*

$$\varphi(x) = \lambda \int_a^b K(x, s)\, \varphi(s)\, d\sigma(s) \qquad (3)$$

où $\sigma(s)$ est une fonction croissante, le noyau $K(x, s)$ est continu et non négatif pour $a \leqslant x,\ s \geqslant b$.

Alors, pour l'existence d'une solution positive $\varphi_0(x)$ de l'équation intégrale (3) avec le nombre caractéristique correspondant λ_0, qui est la plus petite en module racine simple et positive de l'équation

$$D(\lambda) = 0 \qquad (4)$$

il faut et il suffit qu'il existe un entier positif $n_0 > 0$ tel que la m-ième itération du noyau

$$K^{(m)}(x, x) > 0 \qquad (a \leqslant x \leqslant b)$$

pour toutes les valeurs entières de m vérifiant l'inégalité $m \geqslant n_0$.

Pour démontrer le théorème 2 nous aurons besoin du lemme suivant.

Lemme. Considérons une matrice non négative

$$A = \| a_{ij} \| \qquad (i,\ j = \overline{1, n})$$

telle que $a_{ij} \geqslant 0$ pour $|i-j| > 1$, et $a_{ij} > 0$ pour $|i-j| \leqslant 1$. Alors l'équation caractéristique qui lui correspond

$$A(\lambda) = |E\lambda - A| = 0$$

possède une racine simple positive dont le module surpasse les modules de toutes les autres racines de cette équation.

Comme dans les conditions du lemme la matrice A est non décomposable et primitive, ce lemme est une conséquence immédiate d'un théorème de Romanovsky [5]. La notion de matrice non décomposable et primitive a le sens que lui a attribué Frobenius [6].

Démonstration du théorème 2. Suffisance. Considérons avec l'équation (3) l'équation (5)

$$\varphi(x) = \mu \int_a^b K^{(m)}(x, s)\, \varphi(s)\, d\sigma(s) \qquad (5)$$

---

*Jentzsch [2] a démontré la suffisance des conditions du théorème en supposant que $K(x, s) > 0$ pour $a \leqslant x,\ s \leqslant b$, et Gantmacher [3] en supposant que $K(x, s) \geqslant 0$, $K(x, x) > 0$ pour $a < x,\ s < b$. Ils conservent aussi l'hypothèse que le noyau est continu.

où $m \geqslant n_0$; en vertu des équations du théorème 2 et du théorème de Gantmacher elle possède une solution positive $\varphi_0(x)$ et le nombre caractérisique correspondant μ_0, qui est une racine simple et positive de l'équation de Fredholm

$$D^{[m]}(\mu) = 0 \tag{6}$$

qui correspond à l'équation (5), et d'ailleurs le module de μ_0 est inférieur au module de toutes les autres racines de l'équation (6).

Nous démontrerons maintenant que parmi les valeurs de $\sqrt[m]{\mu_0}$ c'est seulement la valeur positive arithmétique que nous désignerons par λ_0 qui sera la racine de l'équation (4). Il est alors évident que λ_0 sera une racine simple, et $\varphi_0(x)$ une solution positive de l'équation (3) correspondant au nombre λ_0.

En effet, en tenant compte de (3) construisons un système d'équations algébriques linéaires, ainsi qu'on le fait ordinairement (4)

$$\varphi_j = \lambda \sum_{k=1}^{n} K_{jk} \varphi_k \sigma(\Delta_k) \quad (j = \overline{1, n}) \tag{7}$$

où $\sigma(\Delta_k) = \sigma(x_{k+1}) - \sigma(x_k)$, et d'ailleurs nous conservons les notations de Goursat.

Ensuite formons la matrice correspondant au système (7)

$$\Phi_n = \| K_{ij} \sigma(\Delta_j) \| \quad (i, j = \overline{1, n})$$

et écrivons sa m-ième itération ($m \geqslant n_0$)

$$\Phi_n^m = \| K_{ij}^{(m)} \sigma(\Delta_j) \| \quad (i, j = \overline{1, n})$$

Il est évident que dans l'équation caractéristique

$$D_n^{[m]}(\mu) = 0 \tag{8}$$

correspondant à la matrice Φ_n^m le premier membre $D_n^{[m]}(\mu)$ tend pour $n \to \infty$ vers une fonction entière $D^{[m]}(\mu)$, si le maximum des intervalles figurant dans la construction du système (7) tend vers zéro.

Choisissons maintenant le maximum des intervalles assez petit, donc le nombre n assez grand pour que $K_{jj}^{(m)}$, $K_{j,j+1}^{(m)}$ et $K_{j-1}^{(m)}$ soient positifs pour tous les $j = \overline{1, n}$. En vertu des conditions du théorème il est facile d'obtenir ces inégalités quand $m \geqslant n_0$. Alors, en vertu du lemme démontré, la matrice Φ_n sera non décomposable et primitive, et sa racine positive λ_n^0, qui existe et dont le module est inférieur à celui de toutes les autres racines, doit vérifier l'inégalité (5)

$$\frac{1}{\max\limits_i \sum\limits_j K_{ij}\Delta_j} \leqslant \lambda_n^0 \leqslant \frac{1}{\min\limits_i \sum\limits_j K_{ij}\Delta_j} \quad (i, j = \overline{1, n}) \tag{9}$$

En passant à la limite dans (9) pour $n \to \infty$ et $\max\limits_{1 \leqslant j \leqslant n} \Delta_j = \Delta^{(n)} \to 0$, nous aurons

$$\frac{1}{\sup\limits_x \int\limits_a^b K(x, s)\, d\sigma(s)} \leqslant \lambda_0 \leqslant \frac{1}{\inf\limits_x \int\limits_a^b K(x, s)\, d\sigma(s)} \tag{10}$$

où $\lambda_0 = \lim\limits_{n \to \infty} \lambda_n^0$ et λ est une racine de l'équation $D(\lambda) = 0$.

En tenant compte de ce que Φ_n est primitive et le noyau est continu sur $[a, b]$ il est facile de voir que

$$\inf_x \int_a^b K(x, s)\, d\sigma(s) > 0$$

Donc, d'une part la fonction $D_n(\lambda) = D_n^{(1)}(\lambda)$ possède, à partir d'une certaine valeur de n et pour toutes les valeurs suivantes, une racine simple λ_0 dont le module est inférieur à ceux de toutes les autres racines et qui vérifie les inégalités (9) et, à la limite, les inégalités (10); d'autre part, la fonction $D(\lambda)$, qui est la limite de $D_n(\lambda)$, possède des racines parmi les valeurs de $\sqrt[m]{\mu_0}$ et ces racines sont simples; on peut en conclure que l'équation $D(\lambda) = 0$ possède une racine simple positive λ_0, qui est inférieure aux modules de toutes les autres racines.

Nécessité. Soit λ_0 une racine simple positive de l'équation $D(\lambda) = 0$, et soit λ_0 inférieur aux modules de toutes les autres racines de $D(\lambda)$. Démontrons alors que pour le noyau continu et non négatif de l'équation (3) il existe un entier positif m_0 tel que $K^{(m)}(x, x) > 0$ $(a \leqslant x \leqslant b)$, quand $m \geqslant m_0$.

En effet, prenons n assez grand pour que l'équation $D_n(\lambda) = 0$, qui est l'équation caractéristique pour la matrice Φ_n, possède une racine simple positive qui soit inférieure aux modules de toutes les autres racines, ce qui est possible parce que

$$\lim_{n \to \infty} D_n(\lambda) = D(\lambda)$$

Alors la matrice Φ_n sera non décomposable et primitive, donc il existe un m_0 tel que $K_{ij}^{(m)} > 0$ pour tous les $m \geqslant n_0$, et $j = \overline{1, n}$. Ensuite, en choisissant convenablement le nombre n, donc les quantités Δ_j $(j = \overline{1, n})$, et en tenant compte de ce que le noyau est continu, il est facile de conclure que l'on a

$$K^{(m)}(x, x) > 0 \quad \text{pour tous les } m \geqslant m_0 \text{ et } a \leqslant x \leqslant b$$

ce qu'il fallait démontrer.

Au cours de la démonstration du théorème précédent nous avons établi la proposition suivante.

Théorème 3. *Le nombre caractéristique primitif de l'équation intégrale* (3) *vérifie les inégalités*.

$$\frac{1}{\sup \int_a^b K(x, s)\, d\sigma(s)} \leqslant \lambda_0 \leqslant \frac{1}{\inf \int_a^b K(x, s)\, d\sigma(s)}$$

où

$$\inf \int_a^b K(x, s)\, d\sigma(s) > 0.$$

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LITTÉRATURE CITÉE

[1] M. Fréchet, Bull. Soc. Math. de France, LXII, fasc. I—II, 68 (1934). [2] R. Jentzsch, J. f. reine u. ang. Math., 141, H. 4, 235 (1912). [3] F. Gantmacher, C. R. Acad. Sci. URSS, I (X), No. 1 (1936). [4] Э. Гурса, Курс математ. анализа. [5] V. Romanovsky, Acta Math., 66, 147 (1931), [6] G. Frobenius, Sitzungsber. d. Akad. d. Wissensch. Berlin, 456 (1912).

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MATHÉMATIQUES

SUR LES ZÉROS DES *L*-FONCTIONS DE DIRICHLET

Par N. TCHUDAKOV

(Présenté par I. M. Vinogradov, de l'Académie, le 14.VI.1945)

1. Dans mes travaux précédents [1] j'ai indiqué une borne supérieure et une borne inférieure pour les zéros de la fonction $L(s, \chi)$ dans la bande critique $0 < \sigma < 1/2$. La connaissance de ces bornes a permis d'améliorer l'évaluation du reste dans la représentation asymptotique du nombre de nombres premiers $\pi(x, k, l)$ dans une progression $l + kv$, où $(k, l) = 1$. Les mêmes résultats ont permis d'améliorer l'évaluation de la quantité

$$\psi(x+u, k, l) - \psi(x, k, l)$$

où $\psi(x, k, l)$ est la fonction de Tchebychev, et $u \leqslant x$.

Dans la présente Note j'expose brièvement les généralisations qui ont été obtenues dans ces problèmes.

Notations. k et l désignent des entiers positifs, $(k,l) = 1$, $x > 0$, $C_1, C_2, \ldots$ sont des constantes absolues. Les symboles «O» et «$\ll$» remplacent des constantes absolues ou bien des constantes qui ne dépendent que d'un paramètre $A > 0$.

Théorème 1. *Soit $L(s, \chi)$ une fonction quelconque de Dirichlet appartenant à un module $k \geqslant 3$; il existe un nombre $C_1 > 0$ tel que $L(s, \chi) \neq 0$ dans le domaine défini par les conditions*

$$0 < |t| \leqslant t_0, \quad \sigma \geqslant 1 - C_1 (\lg k)^{-1}$$
$$|t| > t_0, \quad \sigma \geqslant 1 - C_1 (\lg |t|)^{-\gamma}$$

où $\lg t_0 = (\lg k)^{1/\gamma}$, $0 < \gamma < 1$, γ est une constante absolue.

La quantité γ dépend de l'état actuel des méthodes pour évaluer les sommes de Weyl [2]. La démonstration du théorème 1 est basée sur la synthèse de la méthode de Vinogradov et celle de Landau-Page [3].

Des méthodes analogues aux méthodes classiques permettent de déduire du théorème 1 le

Théorème 2. 1. *Soit $1 \leqslant k \leqslant \exp(\lg x)^{\frac{\gamma}{1+\gamma}}$; alors*

$$\pi(x, k, l) = h^{-1} \sum_{2 \leqslant n \leqslant x} \frac{1}{\lg n} - E \frac{\chi(\bar{l})}{h} \sum_{2 \leqslant n \leqslant x} \frac{n^{\beta_1 - 1}}{\lg n} + O\left(x e^{-C_2 (\lg x)^{\mu}}\right)$$

où $h = \varphi(k)$, $E = 0$ ou 1 et ne dépend que de k; χ_1 est le caractère «exceptionnel» pour tous les $k' \leqslant k$; β_1 est le zéro «exceptionnel» de $L(s, \chi_1)$

$$\mu = \frac{1}{1+\gamma} > \frac{1}{2}$$

Le sens du mot «exceptionnel» est expliqué dans l'article de Page (3).

2. *Soit A un nombre quelconque >0, $k\leqslant(\lg x)^A$; alors*

$$\pi(x, k, l)=h^{-1}\operatorname{li} x+O\left(xe^{-C_2'(\lg x)^{\mu}}\right).$$

3. Le théorème 1 permet aussi d'améliorer les évaluations pour d'autres fonctions arithmétiques, par exemple, pour les fonctions de Tchebychev

$$\psi(x, k, l)=\sum_{\substack{n\equiv l(\mathrm{mod}\,k)\\ n\leqslant x}}\Lambda(n)$$

Théorème 3. *Si $x\to\infty$, on a*

$$\psi(x+u, k, l)-\psi(x, k, l)\sim h^{-1}u$$

où $k\leqslant(\lg x)^A$, $A>0$ est arbitraire, $u=x^{\theta}$, $0<\theta<{}^3/_4+t$.

La démonstration du théorème est basée sur quelques généralisations des formules connues de Mangoldt. A savoir, soit $N(T, \chi)$ le nombre de zéros de la fonction $L(s, \chi)$ dans le rectangle $0<\sigma<1$, $0\leqslant t\leqslant T$, $T>1$; alors

$$N(T, \chi)=\frac{T}{\pi}\lg\frac{Tk}{2\pi}-\frac{T}{\pi}+O(\lg kT) \tag{1}$$

où k est le module de $L(s, \chi)$; χ est un caractère primitif. Ensuite, soit

$$x>0,\ 3\leqslant T\leqslant x,\ kT\leqslant x,\ x\leqslant y\leqslant 2x$$

alors

$$\psi(x, k, l)=yh^{-1}-h^{-1}\sum_{\chi}\bar{\chi}(l)\sum_{|\rho|\leqslant T}\frac{x^{\rho}}{\rho}+O\left(\frac{x\lg^2 x}{T}\right) \tag{2}$$

où $\rho=\beta+\tau i$ parcourt tous les zéros de $L(s, \chi)$ dans le rectangle $0<\sigma<1$, $|\rho|\leqslant T$.

D'ailleurs, on généralise la formule d'Ingham (4)

$$N_{\chi}(\sigma, T)=O((Tk)^{b(1-\sigma)}\lg^{C_4}T) \tag{3}$$

où $N_{\chi}(\sigma, T)$ est le nombre de zéros de $L(s, \chi)$ $L(s, \bar{\chi})$ dans le rectangle ${}^1/_2<\sigma\leqslant\sigma'<1$, $|t|\leqslant T$

$$L({}^1/_2+it, \chi)\leqslant(kT)^C,\ b=2(1+2C)$$

Si l'on choisit maitenant α de manière à avoir

$$0<\alpha<b^{-1},\ \theta>1-\alpha,\ T=x^{\alpha},\ u=x^{\theta},\ k\leqslant(\lg x)^A$$

alors (2) donne

$$\psi(x+u, k, l)-\psi(x, k, l)=$$

$$=uh^{-1}-h^{-1}\sum_{\chi}\bar{\chi}(l)\sum_{|\rho|\leqslant T}\frac{(x+u)^{\rho}x^{\rho}}{\rho}+O\left(\frac{x\lg^2 x}{T}\right)$$

$$\frac{\psi(x+u, k, l)-\psi(x, k, l)}{uh^{-1}}=1+O\left(h\sum_{|\rho|\leqslant T}x^{\beta-1}\right)+O\left(x^{1-\alpha-\theta}(\lg x)^{A+2}\right) \tag{4}$$

En vertu de (1), (3), du théorème 1 et du théorème de Siegel, nous avons

$$\sum_{|\rho|\leqslant T}x^{\beta-1}=2x^{-1}N_{\chi}(0, T)+2\int_0^1 N_{\chi}(\sigma, T)x^{\sigma-1}\lg x\,d\sigma\ll$$

$$\ll x^{-1}T\lg x+\int_0^{1-C_1(\lg T)^{-\gamma}}T^{b(1-\sigma)}x^{\sigma-1}(\lg x)^{C_4+1}\,d\sigma\ll$$

$$\ll o(h^{-1}) + \int_0^{1-C_1(\lg T)^{-\gamma}} x^{(ab-1)(1-\sigma)} (\lg x)^{C_4+1} d\sigma \ll$$

$$\ll o(h^{-1}) + e^{(ab-1)C_1 a^{-\gamma}(\lg x)^{1-\gamma}} (\lg x)^{C_4} + O(x^{ab-1} (\lg x)^{C_4}) = o(h^{-1})$$

Donc, en vertu de (4) on a

$$\psi(x+u, k, l) - \psi(x, k, l) \sim uh^{-1}$$

La méthode classique permet d'en déduire

$$\pi(x+u, k, l) - \pi(x, k, l) \sim \frac{x}{h \lg u}$$

si $k \leqslant (\lg x)^A$.

La valeur classique que l'on trouve pour c est $\frac{1}{2}$, mais on peut améliorer cette valeur.

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LITTÉRATURE CITÉE

[1] N. G. Tchudakov, Rec. Math., 1 (43) (1936). [2] И. М. Виноградов, Тр. Математ. ин-та им. В. А. Стеклова, X (1937). [3] A. Page, Proc. London Math. Soc., 39 (1935). [5] A. E. Ingham, The Quarterly Journ. of Mathem., Oxford ser., 8, No. 32 (1937).

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ASTRONOMY

THE MAGNETIC FIELD OF THE SUN SPOTS

By L. E. GUREVICH and A. I. LEBEDINSKY

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Observation reveals that the magnetic feild of a Sun spot has its lines diverging from the Sun's surface as if the spot was a straight magnet with its axis directed along a radius of the Sun. It is natural to suppose that in the interior of the Sun the field has a similar structure. On the axis of a Sun spot, that is on the axis of this magnet, we may therefore suppose the field to be directed along the vertical.

There is no ground whatever to believe that the circulation of the gas in a Sun spot may often assume the character of a cyclone. No doubt, the gas flows in to the axis of symmetry of the Sun spot in some regions, and away from it in the others.

A circulation of this kind in a conductive gas can lead to the formation of a strong magnetic field by self-excitation from an initially weak field, say, from the general magnetic field of the Sun. In fact, the gas of a Sun spot is known to be strongly ionized. It appears therefore that the gas flowing towards the axis of the spot represents a contracting conductor. As it contracts, the conductor is crossed by a decreasing magnetic flux, and a current is induced in it which tends to restore the magnetic flux, thereby increasing the magnetic field. The result of this process is self-excitation of the magnetic field in those regions of the spot where the gas flows to the axis. The self-excitation leads to a quasi-stationary state, independent of the initial field. In the regions where the gas flows away from the axis, we have an expanding conductor, and the current induced in it weakens the external field.

The spontaneous excitation as well as the spontaneous weakening of the field have different values at different levels of a spot on account of the differences in velocity, density and temperature. Observational data on the magnetic field of the spots relate to their upper portions where the field is weaker than at a certain depth in the middle. Both in the regions of self-excitation and in those of spontaneous weakening there appears an electromagnetic force which tends to diminish the gas velocity component perpendicular to the field. Owing to this, the lines of the gas velocity all but coincide with the lines of the magnetic field, and the velocity components perpendicular to the lines of force are very small.

In that region of the spot where the magnetic field is very intense, Hall's current exceeds the conductance current. Here the electromagnetic force creates a circulation of cyclonic nature (the gas rotates about an axis). Its velocity, though small in comparison with the longitudinal velocity (directed along the field), is yet greater than the velocity of the meridional circulation which takes place in the vertical plane crossing the axis of the spot.

With the difference in pressure at the same horizontal level inside and outside the spot, Δp, the magnetic field H is connected by the relation

$$H = \sqrt{8\pi\Delta p} \qquad (1)$$

The equality is approximate and takes place only in the lower and central parts of a Sun spot. In the external layers which alone are accessible to observation,

$$H \gg \sqrt{8\pi\Delta p} \qquad (2)$$

The depth of the spot and the strength of its magnetic field can be estimated if the value of radiation deficiency is known. The deficiency of radiation of a spot is understood to be the difference between the radiation fluxes from a unit area of the Sun surface and a unit area of the spot surface. The numerical value of the radiation deficiency may be taken to be

$$\eta = 3 \cdot 10^{10} \text{ erg} / \text{cm}^2 \text{ sec.}$$

The energy due to radiation deficiency is accumulated gradually during the lifetime of the spot and is expended in the creation of the magnetic field and in the acquirement by the spot of an additional heat and kinetic energy, $\Delta\varepsilon$.

Denoting the surface area of the spot by S, its volume by V, we shall have, by virtue of the law of conservation of energy,

$$\eta S = \left(\frac{H^2}{8\pi} + \Delta\varepsilon\right)\frac{dV}{dt} \qquad (3)$$

It can be proved, first, that the quantities $H^2/8\pi$ and $\Delta\varepsilon$ are of the same order of magnitude, and, secondly, that their sum is also a magnitude of this order. Therefore

$$\frac{H^2}{8\pi} h \approx \eta t$$

where h is the depth in the spot. In the whole volume of the spot the magnetic field comes to be established very slowly. The time constant, which depends on self-inductance and the resistance of the mass of gas permeated by the currents, is of the order of 10^{10} seconds, not less than the lifetime of the spot.

It may therefore be supposed that all the time that the spot is in existence, the magnetic field is spreading in it in the vertical direction. As the depth of the spot h cannot possibly exceed the radius of the Sun R, we have

$$\frac{H^2}{8\pi} > \frac{\eta t}{R} \qquad (4)$$

Hence $H > 5000$ gauss.

We can as well evaluate the lower boundary of the depth of the spot, that is the upper boundary of the field strength.

Obviously, $\Delta p < p$, therefore $H^2 < 8\pi p$. Hence

$$\eta t \approx \int_0^h \frac{H^2}{8\pi} dz < \int_0^h p\, dz \qquad (5)$$

In the photosphere the dependence of the pressure on the depth is determined by the conditions of radial equilibrium. So equation (5) after some transformations changes to

$$h > \left(15\frac{\varkappa}{g}\eta t\right)^{1/5}\left(\frac{4RT_0}{gM}\right)^{4/5} \qquad (6)$$

where $\varkappa$ is coefficient of absorption, g is acceleration of gravity, M is molecular weight, R is gas constant, T_0 is temperature of the Sun's surface. To know the exact value of the absorption coefficient is not very essential, since it enters into

(6) as a $^1/_5$ power. Putting $\eta t = 10^{17}$, $\varkappa/g = 0.01$, $T_0 = 5 \cdot 10^3$, $M = 1$, we get $h > 20000$ km, and accordingly, $H < 30000$ gauss.

The magnetic field arising in the spot by self-induction does not depend for its sign upon the direction or form of the circulation of the gas, and is entirely determined in this respect by the sign of the initial magnetic field. The case is quite analogous to that of a generator with self-excitation. So the bipolar nature of the Sun spots can be explained on the hypothesis of self-excitation. To put it roughly, the explanation is as follows. The magnetic field in the spot is by far stronger than the general magnetic field of the Sun. On this account the magnetic field about the spot is opposite in sign to that inside. This field of opposite sign gives origin by self-excitation to the magnetic fields of the spots newly formed in the region, which are bound therefore to have a reverse polarity.

The absence of granulation in the region of the shadow is accounted for by the action of the electromagnetic force, which is opposed to all transverse movement and favours circulation only along the lines of force.

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ASTRONOMIE

LE NOUVEAU DE LA PLANÈTE MARS

Par G. A. TIKHOV, membre correspondant de l'Académie

J'ai observé la planète Mars à Poulkovo en 1909 par la méthode photographique au moyen d'un réfracteur de 30″ et en 1918—1929 visuellement au moyen d'un réfracteur de 15″.

Les clichés de 1909 ont eté étudiés au cours des années suivantes, et les résultats ont été publiés dans une série d'articles (1). Quant aux observations de 1918 et de 1920, celles-ci n'ont été détaillement étudiées que l'année courante, et j'ai trouvé alors quelques phénomènes, jettant une certaine lumière nouvelle sur la nature de Mars.

Les observations de 1918 et 1920 ont été effectuées par moi aussi bien sans écrans, qu'avec des écrans rouge foncé, rouge clair, jaune, vert et bleu clair. De plus, au moyen d'un spectroscope à oculaire a été observé le spectre des formations distinctes. Les observations sont représentées sur 120 figures munies des remarques.

Ces observations confirment entièrement les déductions faites par moi d'après les clichés de 1909, c'est-à-dire que la meilleure visibilité des dites «mers» et «canaux» a lieu à travers l'écran rouge, la meilleure visibilité des extrémités polaires a lieu à travers l'écran vert, une bonne visibilité des détails du bord du disque de la planète est observée à travers l'écran rouge, ils disparaissent de vue en cas de l'écran vert, etc.

Mais de plus, en 1918 et en 1920 ont été notés de faits nouveaux, qui sont sujet de la description présente.

On observe presque toujours à travers les écrans vert et bleu sur les bords est et ouest du disque de la planète des taches, tandis que l'éclat des extrémités polaires occupe la seconde place. Tout le disque est traversé par des bandes claires, reliant les taches de l'est et de l'ouest. Parfois une bande claire unit l'extrémité polaire à une bande orientée suivant une latitude.

Les bandes claires sont observées aussi bien dans l'hémisphère d'hiver que d'été.

Il est remarquable que les bandes claires vont dans l'hémisphère d'hiver suivant le bord des «mers» et ne les traversent que rarement.

Ces phénomènes peuvent être expliquées sans difficultés par le climat dur de Mars, et par l'apparition du givre le matin et le soir, même dans les regions équatoriales de la planète.

Les bandes claires correspondent alors à la formation du givre ou bien à l'apparition des nuages. Cela se confirme par le fait que ces bandes apparaissent parfois sur un fond foncé du côté de diminution de la phase.

Mais pourquoi alors les bandes claires évitent-elles les «mers»? Cela peut se produire du fait qu'une part du givre ou de la neige soit cachée par des plantes, couvrant les «mers». Pour cette raison il faudrait considérer les «mers» plutôt comme couverture végétative, ou tout simplement, comme prairie.

A la venue de 13 mai 1920 j'ai noté que les prairies du sud, à travers l'écran jaune, paraissaient être verdâtres, et celles du nord—brunâtres.

A cette époque dans l'hémisphère sud de Mars était le milieu d'hiver, et dans celui du nord—le milieu d'été. A la première vue on peut voir là une contradiction. Or, si l'on suppose que dans les prairies poussent les plantes qui changent leurs couleurs à l'arrivée de l'automne, à côté des herbes et des arbres qui restent constamment verts, cette contradiction se résoud d'une façon très simple.

Mais pourquoi alors observe-t-on le teint brunâtre déjà au milieu d'été? Probablement, parce que le climat de Mars est très sec, et que les plantes, qui brunissent dans les zones tempérées de la Terre seulement à l'arrivée de l'automne, brunissent sur le Mars déjà au milieu d'été. Ce phénomène est tout analogue à celui qui a lieu, par exemple, dans les steppes de l'Asie moyenne.

Pendant l'observation de 13 mai 1920 sans écran il a été noté que l'hémisphère du nord est brunâtre et celui de sud—verdâtre. Cela confirme, par conséquent, les observations avec l'écran jaune.

Enfin, la même chose est confirmée par ce que le 13 mai, lors de l'observation à travers l'écran vert, la couverture végétale était plus foncée dans l'hémisphère du nord que dans celui du sud.

On a donc des raisons à croire que sur le Mars existe la végétation qui perd sa couleur verte au milieu d'été, aussi bien que la végétation constamment verte, tout à fait du type terrestre, c'est-à-dire la végétation à aiguilles et des mousses.

Il y a tout de même un phénomène, qui me paraissait avant incompatible avec l'admission de l'existence sur le Mars de la végétation du type terrestre: c'est que la végétation verte disperse très fortement des rayons infra-rouges. Il semblait que la végétation de Mars devait de même posséder cette propriété. Cependant les observations photographiques de Charonov à Tachkent[2] n'avaient point décelé ce phénomène. Le moyen d'expliquer cette contradiction fut suggéré par l'agrométhéorologue A. Koutyreva (Alma-Ata). Elle attira mon attention sur ce que pendant leur adaptation au climat dur et sec de Mars, les plantes pouvaient évoluer dans le sens de diminution de leur pouvoir réflecteur en rayons infra-rouges.

Une telle conclusion résulte des observations de A. Koutyreva montrant que les propriétés de radiation des plantes sont fonctions des conditions méthéorologiques de leur croissance.

En effet, il est très désavantageux pour une plante, poussant dans un climat dur, de réflecter fortement des rayons infra-rouges, portant une énergie considérable.

Étant de même avis sur cette question, je décidai de comparer la radiation des rayons infra-rouges par des plantes à feuilles et à aiguilles en utilisant les observations de E. Krinov. On pouvait s'attendre que le pouvoir réflecteur en rayons infra-rouges est considérablement moindre chez les plantes à aiguilles que chez celles à feuilles. Cette attente s'était entièrement confirmée, ce que nous montrons dans la table qui suit.

Le pouvoir réflecteur

Plante	Longueur d'onde, mμ								
	400	450	500	550	600	650	700	740	800
Boileau	0.06	0.06	0.10	0.17	0.16	0.14	0.20	9.33	0.54
Sapin	.05	.06	.07	.11	.10	.09	.10	.16	.16
Avoine	.06	.05	.04	.10	.06	.04	.15	.75	.82
Genièvre (en toundra)	.03	.04	.04	.10	.08	.07	.12	.24	.32

Nous voyons qu'en cas des mêmes valeurs pour le boileau et le sapin dans les rayons bleus (450 mμ) le pouvoir réflecteur du boileau en rayons infra-rouges (800 mμ) est trois fois plus grand que celui du sapin.

Pareillement pour les mêmes valeurs correspondant à l'avoine et au genièvre de la toundra en rayons verts (550 m μ), le pouvoir réflecteur de l'avoine en rayons rouges (740 m μ) est trois fois plus grand que celui du genièvre.

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[1] G. A. T i k h o v, Bull. Acad. Sci. URSS, 1039 (1909); 881 (1910); Изв. Пулк. обсерв., № 42, 73 (1911); Изв. Русск. астр. об-ва, XVII, № 5, 169 (1911); Природа, VI, 755 (1912). [2] В. В. Ш а р о н о в, Астроном. журнал (1940); Циркуляры ГАО, 1941.

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GEOPHYSICS

ON THE ORIGIN OF TERRESTRIAL MAGNETISM

By J. FRENKEL, Corresponding Member of the Academy

§ 1. General Principles. The explanation of the origin of Terrestrial Magnetism proposed in this paper is a product of a combination of ideas put forward by several different authors which singly failed to yield the explanation required. These ideas may briefly be stated as follows.

1. The central core of the Earth with a radius $R=3000$ km consists of a molten metal mass with a relatively low viscosity. This may be inferred from seismic data which show that transversal waves penetrate to a depth of 3500 km and are not propagated through the Earth's core. The reason lies in the fact that the viscosity of molten metals is but very slightly affected by pressure ([1]).

2. The existence of a sufficiently large temperature gradient in the Earth's core must give rise to convection currents in the liquid of which it is made ([1,2]).

3. Owing to the high electric conductivity of this liquid, its motion in some initially weak magnetic field of adventitious origin can be accompanied by a gradual increase of this field by means of a mechanism, similar to the self-excitation of a dynamo. This idea has recently been used by L. Gurevich and A. Lebedinsky in explaning the magnetic field of Sun spots on the assumption that the rôle of the initial («residual») field is played by the general field of the Sun ([3]).

In the author's opinion this general field of the Sun may be due to the same cause which accounts for the origin of the geomagnetie field, namely, the convective motion of the electrically conducting mass within the Sun.

§ 2. Temperature Distribution in the Earth's Core. If the Earth is assumed to have reached a steady state, the distribution of the temperature within it can be determined by the equation $\frac{\varkappa}{r}\frac{d^2(rT)}{dr^2}+q=0$, where r is the distance from the centre, $\varkappa$ is the normal conductivity, and q is the heat generated per unit time and volume (by radioactive substances); for the sake of simplicity $\varkappa$ and q will be treated as constants within the Earth's core. The solution of this equation is $T=T_0-qr^2/6\varkappa$, where T_0 is temperature at the centre. The temperature gradient $-dT/dr=qr/3\varkappa$ is thus proportional to the distance r, and consequently to the acceleration of gravity $w=\frac{g}{R}r$ (g being the value of w at the surface of the core, and R, the radius of the latter). The condition of stability of the liquid core with respect to free convection is expressed by the inequality $-dT/dr<w\alpha T/c_p$, the right-hand side of which represents the adiabatic value of $-dT/dr$ (α is the coefficient of thermal expansion, c_p is specific heat at constant pressure). So long as the variation of T within the core is relatively small, the preceding condition can be rewritten in the form $q/3\varkappa<g\alpha T_0/Rc_p$, and is thus seen to be satisfied for all values of $r\leqslant R$. We see that if $q>q_0=3a\varkappa T_0\ g/R$,

where $a = \varkappa / c_p$, the liquid core of the Earth must be in a state of free convective motion. Putting here $a = 1$, $\alpha = 10^{-5}$, $T_0 = 3 \cdot 10^3$ and $g/R = 10^{-5}$, we get $q_0 = 10^{-7}$ erg/cm³ sec. It should be mentioned for the sake of comparison that the heat generated in the external portions of the Earth's crust by radioactive substances is equal to 10^2 erg/cm²sec.

§ 3. Convective Motion in the Earth's Core. In the presence of convection currents the temperature distribution must, of course, be different from the one considered above. In the region of ascending currents the temperature must be higher by a certain amount ΔT_h than in the region of descending currents at the same level. This circumstance follows from the examination of the convective circulation in a rectangular tube representing schematically a closed tube of flow (a «jet»). If the circulation of the liquid is barred by a diaphragm in one of the horizontal portions of the tube, this diaphragm must suffer a pressure $\Delta p = \Delta \rho_h \bar{w} R$, where R is the height of the vertical sides of the tube (which can be identified with the radius of the core), $\bar{w}$ is the average value of the acceleration, and $\Delta \rho_h$ is the average difference of the densities of the liquid at the same level in the two vertical portions of the tube. This pressure

$$\Delta p = R \bar{w} \rho \alpha \Delta T_h \qquad (1)$$

is the mechanical analogon of the electromotive force in an open thermoelectric circuit. If the diaphragm is removed, the liquid must circulate in the tube, the work done per unit time being equal to the product of Δp and the corresponding volume flow (which plays the rôle of the current strength). Denoting the total volume of the liquid in the tube with V and the period of its complete resolution with τ, we can represent this flow by the ratio V/τ, so that the power of the «thermo-gravitational» cell is equal to

$$E = \Delta p \cdot V/\tau. \qquad (2)$$

If V is understood to represent the total volume of the circulating liquid core, then this expression gives the total power of the «thermo-gravitational» engine which this core constitutes.

This power must be equal to the total heat qV generated in the core per unit time, multiplied by the efficiency η which, in its order of magnitude at least, is equal to the ratio of ΔT_v to T_0 where ΔT_v is the difference between the temperature at the centre and at the surface of the core.

We thus get $\Delta p \cdot V/\tau = qV \Delta T_v / T_0$, that is, according to (1)

$$\tau \frac{\Delta T_v}{\Delta T_h} = \frac{R \bar{w} \rho \alpha T_0}{q} \qquad (3)$$

If the resistance to the circulation of the liquid was due solely to its viscosity, the power (2) should be equal to the product of V with the viscosity coefficient μ and the square of the velocity gradient, which must be of the same order of magnitude as $1/\tau$. We should thus have in this case $\Delta p V/\tau = V\mu/\tau^2$, *i. e.* $\tau = \mu/\Delta p$. Putting here $\mu = 10^{-2}$ poises, we obtain even for absurdly small values of Δp (say, 1 dyne/cm²) a velocity R/τ larger than that of light.

§ 4. Electromagnetic Self-excitation and Resistance Associated with the Free Convection of an Electrically Conducting Liquid. As has already been pointed out in § 1, the circulation of an electrically conducting liquid (such as the Earth's metallic core) must be accompanied by a gradual excitation of a magnetic field (starting from an initially very small adventitious field) and of a system of electric currents producing it. As a result, in addition to the mechanical friction, electromagnetic friction forces must arise tending to slow down the

motion of the liquid, just as in the case of an ordinary self-excited dynamo. It can easily be seen that the latter effect is much more important than the former one, so that in the steady state practically the whole power of the thermo-gravitational engine is converted into Joule's heat of the dynamo which is driven by it (the mechanical circulation representing the former, and the electrical, the latter).

This Joule's heat W is equal to Vi^2/σ, where i is the mean value of the electric current density, and σ is the electrical conductivity of the Earth's core (i and σ will be measured in electrostatic units). On the other hand, according to the equation $\operatorname{rot} H = 4\pi i/c$ (where c is the velocity of light), i must be connected with the average value of the field H by the relation $i/c \cong H/R$. We thus get

$$W \cong VH^2c^2/R^2\sigma \tag{4}$$

Since, further, $i \cong \sigma v H/c$, where $v = R/\tau$ is the approximate value of the electric induction field (so long as the velocity v is not nearly parallel to H) we find replacing i by Hc/R

$$\frac{1}{\tau} = \frac{c^2}{R^2\sigma} \tag{5}$$

From (4) and (5) there follows

$$W \cong VH^2/\tau \tag{6}$$

Equating this to the expression (2) (which corresponds to a neglect of the effect of mechanical viscosity), we obtain

$$H^2 \cong \Delta p \tag{7}$$

This relation * can be obtained directly from a consideration of the equilibrium between hydrostatic and electromagnetic forces, using Maxwell's tensor of magnetic tensions and pressures.

Putting in (5) $c = 3 \cdot 10^{10}$, $R = 3 \cdot 10^8$ cm, and $\sigma = 10^{16}$ (which is approximately the electrical conductivity of iron in ESU), we get $\tau = 10^{12}$ sec. This corresponds to a velocity of motion R/τ of the order of 10^{-4} cm/sec.

§ 5. Magnitude of the Earth's Magnetic Field. For the calculation of the average intensity of the magnetic field in the Earth's core, one can use equation (1) in connexion with (3) and the expression for ΔT_v as a function of q, which follows from the theory of heat transmission in the case of free convection within a hollow sphere. According to Nusselt[5] on the assumption of a constant value and direction of the acceleration of gravity w, the coefficient of heat transfer by convection is given by the formula $A = \frac{\varkappa}{R} K(\mathrm{Gr})^{1/3}$, where $\mathrm{Gr} = \frac{1}{\mu^2} R^3\rho^2 w\alpha\Delta T_v$ is Grashof's number, and K is a numerical coefficient (equal to 0.8 if A is expressed in cal per m^2 per sec.).

Putting $4\pi R^2 A = \frac{4\pi}{3} R^3 q$, we get $A = {}^1/_3 Rq$, so that $q = (3\varkappa/R^2) K(\mathrm{Gr})^{1/3}$, *i. e.* $(R^3/\mu^2)\rho^2 w\alpha\Delta T_v = (R^2 q/3\varkappa K)^3$. Hence

$$\Delta p = \frac{q^4}{T_0}\left(\frac{R}{3\varkappa K}\right)^3 \frac{\mu^2}{\rho^2 w\alpha} \tag{8}$$

We shall not dwell on the application of this formula, but will estimate the order of magnitude of Δp, and consequently of H, with the help of equation (1), putting $\Delta T_h = 1° - 100°$C (higher values being hardly admissible,

* Which has been derived in 1944 by Alfven[4], and independently by Gurevich and Lebedinsky[3] in connexion with Sun spots.

since T_0 does not exceed 3000°K). With $R = 3 \cdot 10^8$ cm, $w = 10^3$ cm/sec^2, and $\alpha = 10^{-5}$, we get $\Delta p = 10^7 - 10^9$ dyne/cm^2, and consequently, $H = 10^4$ oerstedt.

This value exceeds by a factor of 10^4 that which is actually observed on the Earth's surface. It can be lowered by a factor of 10, if one takes into account the fact that the strength of a field connected with a dipole moment decreases with the distance from the centre as $1/r^3$. A further decrease of H is obtained if it is assumed that the convective motion of the Earth's core is split up into a large number N of cells whose linear dimensions l are small compared to its radius R ($l \cong R/N^{1/3}$).

If, moreover, the magnetic moments of the electric currents in the different cells have different directions, with a more or less random distribution, the resulting magnetic field is further decreased in the ratio $1/\sqrt{N} : 1$. It must be emphasized that all the theories of Terrestrial Magnetism which have been proposed thus far, have lead to values of H which are extremely small compared to the observed one (of the order of 10^{-15}).

The theory proposed above is, of course, very incomplete and must further be developed with allowance for the influence of the Coriolis forces due to the rotation of the Earth (which is probably responsible for the approximate parallelism between its rotation and magnetic axis), and with a fuller account of the influence of the modification of the convection of the liquid core as a result of the action of the electromagnetic forces. It should be mentioned that if this action was neglected, then the convection motion would lead to an exponential increase of the initial field with the time at some places and to an exponential decrease at others.

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REFERENCES

[1] D. R. Inglis, Phys. Rev., **58**, No. 6, 577 (1940). [2] W. M. Elsasser, Phys. Rev., **56**, No. 8, 855 (1939). [4] L. E. Gurevich and A. I. Lebedinsky. C. R. Acad. Sci. URSS, XLIX, No. 2 (1945). [5] Alfven, Arkiv för Mathem., Astron. o. Physik (1944). [5] W. Nusselt, Wärmeüb. i. d. Verbr. Kraftm., Mitteil. über Forschungsarbeiten, H. 264. S. 33.

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OCEANOLOGY

HORIZONTAL CIRCULATION OF WATER IN THE SEA, AS CAUSED BY NON-UNIFORM VELOCITY DISTRIBUTION IN A SHOREWARD WIND ACROSS ITS MOVEMENT

By W. B. STOCKMANN

(Communicated by P. P. Shirshov, Member of the Academy, 30.XI. 1944)

In my earlier paper(1) on the stationary circulation brought about by a wind in a closed homogeneous sea I was able to show by an approximate analysis of the phenomenon that non-uniformity in the transverse distribution of velocity of a wind blowing in one direction may give rise to a circulation of water not only in the vertical, but also in the horizontal plane. Therefore, compensation currents directed opposite to the wind may exist at the surface of a sea.

In the present paper some considerations are put forward concerning the rôle (in my opinion, a very important one) which the transverse non-uniformity of a shoreward wind plays in the horizontal circulation observed in closed shallow seas, large bays and in the equatorial regions of the ocean.

To solve this problem in first approximation, the inertia terms and the effect of the Earth rotation will be neglected in the equations of motion, and, omitting, as before, the conditions at the lateral boundaries, only the longitudinal (along the movement of the wind) velocity components of a current in a basin of constant depth H will be considered. Seeing that under real conditions the horizontal dimensions of sea basins by far exceed their depth, the vertical component of the current can (2) be neglected practically on the whole extent of the basin, and the stationary motion in the direction of the wind can roughly be described by a single equation

$$\mu \frac{\partial^2 u}{\partial z^2} + g\rho \operatorname{tg} \gamma = 0 \qquad (1)$$

which is a simplification of the Stokes-Navier equations. Here u is the velocity component along the horizontal axis X (positive is the direction along the wind); g is the acceleration of gravity; ρ is a constant density of water; γ is the inclination angle of the free water surface (we take it to be a plane), created by the wind in the basin as it piles up water in the direction of its movement; z is the vertical coordinate measured downward from the water surface; μ is the coefficient of turbulent viscosity, supposed at first to be constant. It should be noted that the transverse inclination of the water level is neglected, and μ is considered to be invariable with the transverse coordinate y.

The solution of (1) is subject to boundary conditions

$$\mu \left(\frac{\partial u}{\partial z} \right)_{z=0} = -T(y); \quad (u)_{z=H} = 0 \qquad (2)$$

where $T(y)$ is the tangential pressure of the wind at the water surface varying in the transverse direction according to an arbitrary law.

Under these conditions the solution of (1) transforms into

$$u = \frac{g\rho \operatorname{tg}\gamma}{2\mu}(H^2 - z^2) + \frac{T(y)}{\mu}(H - z) \qquad (3)$$

As before, the constant quantity $\operatorname{tg}\gamma$ is determined from the condition that in the stationary case the total discharge of water through the cross-section of the basin is zero, that is

$$\int_0^L \int_0^H u\, dz\, dy = 0 \qquad (4)$$

where L is the horizontal extension of the basin measured across the direction of the wind. Substituting (3) in (4) we get

$$\operatorname{tg}\gamma = -\frac{3}{2g\rho H}\frac{1}{L}\int_0^L T(y)\, dy \qquad (5)$$

Here we have

$$\frac{1}{L}\int_0^L T(y)\, dy = \overline{T}$$

where $\overline{T}$ is the tangential pressure of the wind, averaged over a section of the basin transverse to the wind. Therefore (3) can be represented in the form

$$u = \frac{(H - z)}{\mu}\left[T(y) - \frac{3\overline{T}(H + z)}{4H}\right] \qquad (6)$$

from which it follows that a counter-current ($u < 0$) at the surface ($z = 0$) is possible under the condition that

$$\overline{T}/T \geqslant 1.33 \qquad (7)$$

that is, if the tangential pressure of the wind deviates from the mean (in the direction of smaller T) by more than 25 per cent.

Making use of the generally recognized relation

$$T = \rho' k W^2 \qquad (8)$$

where ρ' it the density of the air, k is a constant coefficient, W^2 is the velocity of the wind, we can rewrite (7) in a form more convenient for practical application

$$\overline{W^2}/W^2 \geqslant 1.33 \qquad (7')$$

Unfortunately, (7′) cannot be expressed directly in wind velocities, because $\overline{W^2} \geqslant (\overline{W})^2$. Nevertheless it is evident that even a very small transverse non-uniformity in the wind velocity distribution can produce at the water surface a compensation current directed opposite to the wind.

It is easy to see that (7) and (7′) do not depend on the absolute value of the coefficient of internal turbulent friction, μ. The magnitude of the above criteria is only influenced by the form of the chosen dependence of μ on z. If, for instance, we chose this dependence in a form approved in seafaring practice as $\mu = \mu_0 \left(\frac{H - z}{H}\right)^{1/2}$, then for estimating the possibility of a counter-current arising at the surface, we have to use the following inequality instead of (7′)

$$\overline{W^2}/W^2 \geqslant 1.20 \qquad (7'')$$

So in this case the possibility for a counter-current to form at the surface is much higher than if $\mu = \text{const}$.

Putting $u=0$ in (6) we can easily obtain the equation of the separation line between shoreward flow and compensation counter-current in the vertical cross-section of the basin. This equation is

$$z=\frac{H}{3\overline{T}}[4T(y)-3\overline{T}] \qquad (9)$$

From (9), it may be remarked, follows the well-known result that in the case of a uniform wind, where $T(y)=\overline{T}=\text{const}$, the separation boundary is a straight line, parallel to Y and running under the water surface at a distance equal to one third of the depth of the basin ($z=H/3$). Another inference to be drawn from (9) is that in the general case the inclination of the separation boundary to the Y-axis at a point (z, y) in the water is defined by the expression

$$\operatorname{tg}\alpha=-\frac{dz}{dy}=-\frac{4H}{3\overline{T}}\frac{dT}{dy}=-\frac{2.67HW}{\overline{W^2}}\frac{dW}{dy} \qquad (10)$$

from which it appears that this boundary at the same point y has an inverse inclination with the curve bringing out the variation of the wind velocity in the transverse direction.

It gives the mirror image of that curve, though on a different scale. As may easily be realized, the quantity $\operatorname{tg}\alpha$ is to a certain extent a qualitative measure of the ratio between the horizontal component v of the water velocity in the basin, directed across the movement of the wind, and the vertical component w.

In the limiting case, where $\operatorname{tg}\alpha=0$, we obviously have $v=0$, and if $\operatorname{tg}\alpha=\infty$, then $w=0$. It follows therefore from (9) that if the mean square velocity of the wind rests invariable, the ratio v/w grows either with the depth of the basin or with the wind velocity and its transverse gradient.

In the vertical section of the basin cut across the direction of the wind the separation boundaries between the currents moving before and toward the wind can be determined speedily by a rapid and simple procedure based on the approximate analysis of the phenomenon, as set forth in this paper.

For this purpose equation (9) may conveniently be rewriten as follows

$$T(y)=\frac{3\overline{T}}{4}\left(1+\frac{z}{H}\right)$$

or, making use of the relation (8)

$$W^2(y)=\frac{3\overline{W^2}}{4}\left(1+\frac{z}{H}\right) \qquad (11)$$

As is easy to see, the equation (11) is satistied for the intersection points between the curve bringing out the transverse variation of squared wind velocities and the straigth lines parallel to Y, whose ordinates are to $\overline{W^2}$ as $\frac{3}{4}\left(1+\frac{z}{H}\right)$ to one. Thus, for determining the possibility of a counter-current at the water surface ($z=0$) and the position of its boundaries, one has to draw in the plot of transverse variation of wind square velocity a straight line parallel to Y and whose points have $\frac{3}{4}\overline{W^2}$ for ordinate. If this line intersects with the square velocity curve (which settles the question as to the possibility of a counter-current at the surface), then the abscissae of the intersection points determine the position of the counter-current boundaries.

The question whether or not these boundaries meet the bottom of the basin, and the determination, if they do, of their position there, is solved by drawing a straigth line with ordinate $\frac{3}{2}\overline{W^2}$ and finding the abscissae of its intersection points wiht the given curve $W^2(y)$. Finally, the outline of the counter-current boundaries at intermediate depths is determined by the projections of the intersection points between straight lines with ordinates $\frac{3}{4}\overline{W^2}(1+0.1)$, $\frac{3}{4}W^2(1+0.2)$ *etc.*, if the depth coordinates are $z=0.1H$, $z=0.2H$, *etc.*, respectively.

A similar graphical method is used in determining the position of the maximum of counter-current velocities. This maximum, upon differentiation of (6), is defined by the equation

$$W^2(y) = 1.5\overline{W^2}\,\frac{z}{H}$$

Fig. 1 is an example of the above-mentioned constructions. The broken line in its upper part brings out the variation of the wind velocity $W(y)$ across the di-

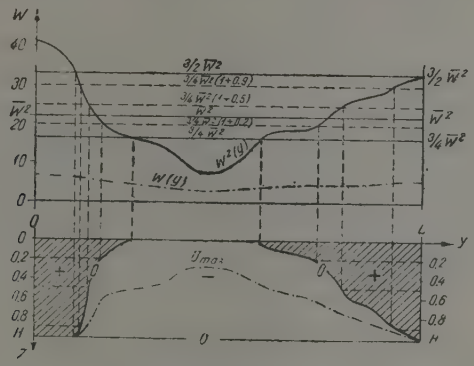

Fig. 1

rection of wind movement which is perpendicular to the drawing. The continuous line represents the variation of the squared velocity of the wind, $W^2(y)$. In the lower part of the figure is indicated the outline of a vertical cross-section of the basin, and are drawn the boundaries between the current of water moving before the wind (+) and the compensation counter-current (—). Finally, the dash- and-dot line shows the position of the counter-current velocity maximum, which in the case considered is always beneath the water surface.

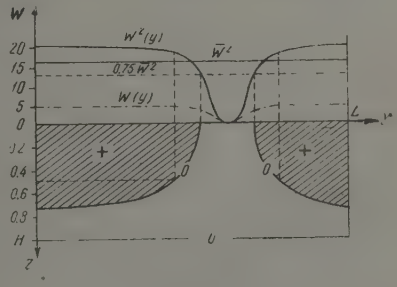

Fig. 2

With equal ease can the counter-current boundaries be found in the case we are particularly interested in, where the velocity of the piling up wind is sharply diminishing in the narrow central region of the basin, while it remains constant beyond this region (Fig. 2). This kind of transverse non-uniformity in a shoreward wind closely resembles the conditions encountered in the equatorial regions of the Pacific. Here the trade winds, which drive water to the western part of the ocean, have a uniform velocity in general, yet the latter drops to zero in the narrow calm belt north of the equator.

Remarkably enough, the shape of the counter-current boundaries, as represented in Fig. 2 (which, according to our theory, is intimately connected with the transverse non-uniformity of the wind), coincides exactly with the outline of the transverse boundaries of the equatorial counter-current in the Pacific (Fig. 3), as drawn by Sverdrup [3] on the basis of hydrological observations.

This appears to shed a new light on the nature of the equatorial counter-currents, which have been proved to arise from the transverse non-uniformity of the trade winds, which pile up water in the equatorial regions of the ocean.

It may be remarked in conclusion that, given the shape of the counter-current boundaries and the mean square velocity of the wind, we can determine the charac-

ter of the transverse non-uniformity of the wind, and for certain areas, to reproduce (by reverse plotting) the transverse profile of the wind velocity. What is more, the above constructions and the formula (10) lead to the conclusion that

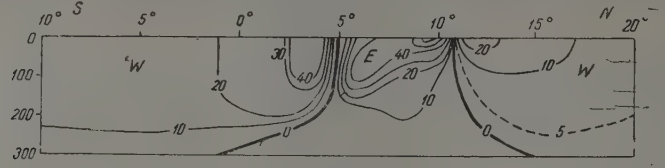

Fig. 3

in the case of a stationary wind the inclination of the counter-current boundary cannot possibly vary in direction with the depth. Otherwise the velocity of the wind should be liable to change in the course of time. So we have here an additional qualitative criterion, and a very simple one, by which to recognize a not stationary wind and the circulation it occasions in the sea.

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REFERENCES

[1] W. B. Stockmann, Bull. Acad. Sci. URSS, sér. géogr. et géophys., No. 1 (1941). [2] K. Hidaka, Mem. of the Imper. Marine Observatory, Kobe, VII, 1 (1939). [3] H. U. Sverdrup, M. W. Jonson, R. H. Fleming, The Oceans, their Physics, Chemistry and General Biology, New York, 1942.

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PHYSICAL CHEMISTRY

ON SHILOV'S THEORY OF CHAIN REACTIONS AND ITS FURTHER DEVELOPMENT

By N. S. AKULOV, Member of the Academy of Sciences of the Belorussian SSR

The fundamentals of the modern theory of simple and branched chain reactions were formulated by Shilov in 1905 (1). To describe the development of these reactions he proposed the following mechanism

$$A' + A = A'', \quad A'' + B = A' + B' \tag{1}$$

where A and B are the reactants, A' and A'', two active intermediate products; and B' is either the end product or the sum of the end and the intermediate products. Thus, at the end of Shilov's cycle there are formed («regenerated») ω_i new molecules of the active product. From a number of examples given by Shilov it is seen that ω_i is either equal to or greater than unity, for example, $\omega_i = 2$. [Later on, in the works of other authors, such reactions came to be known as «branched» reactions (2)]. As a rule, side-by-side with cycles for which ω_i molecules regenerate, there exist cycles for which regeneration does not take place ($\omega_0 = 0$), and this leads to a lowering of the average regeneration ω *.

It is convenient in calculations to employ the expression for a «long chain» in the following form (induction factor) proposed by Shilov:

$$I = \Delta x / \Delta n' = -\Delta x / \Delta n \tag{2}$$

where Δx is the increase in quantity of the end product, or, what amounts to the same thing, the decrease in quantity of the reactant (acceptor)**; and $\Delta n' = -\Delta n$ is the decrease in quantity of the active product (inductor). Thus, $I = \nu$ gives us the «length of a chain», *i. e.* the number of molecules of the end product for each perished and not regenerated molecule of the active product. This formula Shilov (in 1902—1905) and then Bodenstein (in 1913 and later) employed for a quantitative characteristic of chain processes.

Shilov showed that I is positive or negative according as the concentration of active product decreases or increases. Thus, Shilov's induction factor is a conception more general than chain length, and includes the conception of the length of a chain as a special case, when it has a positive value.

Upon N molecules of active product reacting, ωN molecules of it are generated, and at the same time σN molecules of end product appear. Thus, according to (2)

$$I = \frac{\sigma}{1 - \omega} \tag{3}$$

* See, for example, Harkur and Esson, referred to in (1).

** $I = \Delta C_{\text{acc}} / \Delta C_{\text{ind}}$.

The highly important condition of «isothermal» (sometimes called chain) explosion laid down by Shilov

$$I = \infty \qquad (4)$$

must be interpreted to mean that on transition to «isothermal» explosion the active product is regenerated completely ($\omega = 1$) in the course of the process, and so Δn is equal to zero. In attempts to apply Shilov's theory some authors determined his I by means of assumptions introduced *ad hoc* and, as a rule, incompatible with one another. At the same time, it is easy to derive a general equation for computing I.

Suppose that z_i reactions are taking place in a system every second, in the course of which each molecule of active product upon reacting gives ω_i regenerated molecules and σ_i end molecules. Then in the time interval Δt

$$\Delta x = (\sigma_1 z_1 + \sigma_2 z_2 + \ldots)\,\Delta t, \quad \Delta n = [(\omega_1 - 1)\,z_1 + (\omega_2 - 1)\,z_2 + \ldots]\,\Delta t \qquad (5)$$

And so, according to (2),

$$I = \frac{\sum\limits_i \sigma_i z_i}{\sum\limits_i (1 - \omega_i)\,z_i} = \frac{\sum\limits_i \sigma_i \delta_i}{\sum\limits_i (1 - \omega_i)\,\delta_i} \qquad (6)$$

where $\delta_i = z_i \Big/ \sum\limits_i z_i$, and $i = 0, 1, 2 \ldots$

Applying this formula solves the problem of the theoretical computation of the factor I*. The latter is obviously a fundamental problem in the theory of chain processes and the theory of combustion (z_i is determined on the basis of the law of mass action; in the simplest case $z_m = k_m c^m$, where c is the concentration of the initial reactant, and $m = 0, 1, 2 \ldots$).

As for the kinetics of the development of chain processes, Shilov showed that if I is negative we have a continuous increase in the reaction rate over a certain period of time, and the appearance of an induction period τ. Such a type of development of the process Shilov called isothermal explosion; later some authors began to call it «chain degenerated explosion».

The rate of increase in the end product for this case Shilov expressed by the equation

$$dx/dt = k_0 n\,(c_0 - x) \qquad (7)$$

From this (when $I = \text{const}$) Shilov gets for a characteristic of the chain process an equation of the autocatalytic type, which, according to (2), may be written in the following form

$$\frac{dx}{dt} = k_0 \left(n_0 - \frac{1}{I_0}\,x\right)(c_0 - x) \qquad (8)$$

where $I_0 = I_{t=0} < 0$; $n_0 = n_{t=0}$. This kinetic equation he exhaustively checked by analysis of reactions in the liquid phase (which has a number of advantages over checking in the gas phase, for in the latter case thermal auto-acceleration easily distorts the normal isothermal course of the kinetics).

Shilov's equation of type (8) was referred to in some later works as a sort of a «new universal law».

The purpose of the present paper is: 1) to establish the general relation for I, and then to generalize Shilov's equation of kinetics and our equation of autogenesis so as to render them fit to serve as a basis for judgments on the mechanism of the kinetics (and in those cases where this cannot at yet be done, they at least should describe correctly the experimental data available). With this aim in view, let us expand $-1/I$ into a power series with respect to x and, since equation (2) may be written in the form $dn = (-1/I)\,dx$, we get:

$$\frac{dn}{dx} = -\frac{1}{I} = \left(-\frac{1}{I_0}\right)(1 + k_1 x + k_2 x^2 + \ldots) \qquad (9)$$

* In the general case ω_i can by itself be a function of z_j ($j = 1, 2, \ldots$).

From (9), after integration and substitution into (7), we have

$$dx/dt = k_0 (n_0 + b_0 x + b_1 x^2 \ldots)(c_0 - x) \qquad (10)$$

where *

$$b_0 = -1/I, \quad b_1 = \frac{1}{2} b_0 k_1, \ldots$$

The important circumstance which follows as a consequence from this generalized form of Shilov's equation is that if $-1/I$ decreases in the course of the reaction (b_1 is negative), the reaction may practically stop not only owing to the fact that the reactant has been used up, but owing to the ever diminishing regeneration. In this case, side-by-side with the Shilov effect of chain acceleration (chain explosion), there will exist an effect of chain damping when the reacting system passes through the state defined by condition (4).

The presence of such an effect is substantiated by experiments on the oxidation of hydrocarbon and other substances.

However, owing to the evolution of energy in the course of the reaction and in a number of other cases, dn/dt may increase rapidly within a certain interval of time, instead of decreasing. In this case the reaction rate will increase in the initial stage not exponentially, but more rapidly than follows from Shilov's law**. In particular, $-1/I = dn/dx$ may have already received a sufficiently large increment when x has yet received a very small increment (for example, in the case of a practical adiabatic process in the initial stage, especially for rapidly progressing processes). In this case the expansion of $-1/I$ in a power series with respect to x is found to be inconvenient in practice, and in the limiting cases, impossible in principle. Owing to this, it is more expedient to integrate equation (7) in the general form before resorting to expansion of one or another function in series. Upon doing this, we get

$$-\ln(1-\alpha) = k_0 \bar{n} t \qquad (11)$$

where

$$\alpha = \frac{x}{c_0}, \quad \bar{n} = \frac{1}{t} \int_0^t n \, dt \qquad (12)$$

Approximating $\bar{n}$ by a polynomial from α and substituting into (11), we get the «law of autogenesis» that we gave previously in the following generalized form***

$$k_0 t = \frac{\ln(1-\alpha)}{n_0 + n_1 \alpha + n_2 \alpha^2 + \ldots} \qquad (13)$$

Reactions for which such an expansion of $\bar{n}$ into a series in α is possible and which thus obey equation (13) we have termed autogenetic [for (13) makes it possible, in particular, to describe processes with a spontaneously increasing rate].

A comparison with experiment actually shows that a large number of kinetic curves (for example, those obtained by Bone and Hill for the oxidation of ethane) are splendidly described by (13), but cannot be described by (10), the equation of the autocatalytic type. In a large number of cases we may already confine ourselves to the first two terms of the expansion. The family of kinetic curves obtained herewith is plotted in Fig. 1. Experiment gives curves of a similar type. In some cases, however, especially when the first term of the expansion is extremely small and may be neglected, it is necessary to take into account the second and third terms. Taking account of the third term makes it possible, as a rule, to eliminate the deviation that is sometimes observed between the theoretical curves of the

* Having computed I with the help of (6), one may determine b_i.

** The coincidence between this important law and the autocatalytic formula of Ostwald is a purely formal one.

*** In the general case $\bar{n} = \sum_m a_m (\alpha - \alpha)^m$, where $\alpha_0 = \alpha_{t=0}$.

type shown in Fig. 1, and the experimental curves. The rôle of this third term is easy to see from the diagram given in Fig. 2.

It should be noted that when in 1940 I had first derived the formula of autogenesis in a not sufficiently general form, it might have been thought that the fact that the time explicitly enters into it, was a characteristic trait of the equation of autogenesis. Now, however we find directly

$$k_0 t = f(\alpha, n_0, n_1, \ldots)$$

whence by differentiation with respect to α one may always directly obtain

$$d\alpha/dt = k_0 \Phi(\alpha, n_0, n_1, \ldots)$$

i. e. an expression into which the time does not enter explicitly.

In a number of cases it is necessary to take into consideration that the total change in pressure is the sum of the change in pressure due to the change in: 1) the

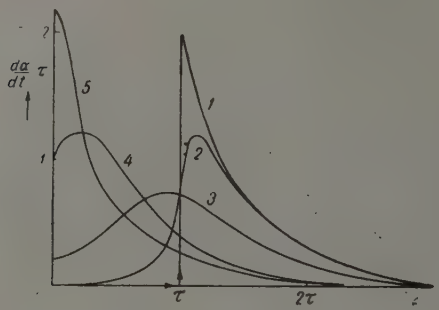

Fig. 1. $a_0 = n_0/n_1$; $\tau = 1/k_0 n_1$; *1*, $a_0 = 0$; *2*, $a_0 = 0{,}01$; *3*, $a_0 = 0{,}2$; *4*, $a_0 = 1{,}0$; *5*, $a_0 = 2.2$; $n_2 = 0$

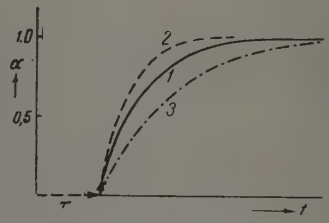

Fig. 2. $n_0 = 0$; *1*, $n_2 = 0$; *2*, $n_2 > 0$; *3*, $n_2 < 0$

number of molecules of the end product, 2) the number of molecules of the intermediate product, and 3) the temperature T. As for the change in temperature in the course of reaction, it may be shown that it may be determined directly from the course of the experimental curves of the kinetics for the pressure, taking into account the heat loss equation and the Klapeyron equation. Afrer this the problem of deriving «isothermal curves» from the observed curves of the kinetics may be solved. Without this, bringing * the curves of the explosive (adiabatic) type into coincidence with the curves of the isothermal kinetics of slow reactions is methodically incorrect.

Still less permissible is displacement (a shift along the t-axis) of the origin of the experimental curves with respect to the origin of the theoretical curves. This comment would be entirely superfluous if such proceedings in applying the Shilov equation had not been encountered in the literature [3].

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4.XII.1944.

REFERENCES

[1] Н. Шилов, Сопряженные реакции окисления, М., 1905. [2] N. Akulov, C. R. Acad. Sci. URSS, XXXIX, No. 5, 6, 8 (1943); XLIII, No. 7 (1944); XLVIII, No. 9 (1945). [3] Н. Н. Семенов, Цепные реакции, 1934, стр. 95.

* By changing the scale along the t-axis.

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PHYSICAL CHEMISTRY

THE PROPERTIES OF ROTATION ISOMERS

By M. V. WOLKENSTEIN

(Communicated by S. I. Vavilov, Member of the Academy, 30. I. 1945)

Today it may be considered a definitely established fact that the so-called internal free rotation around single bonds in molecules is always hindered to some degree or other. A method of allowing for this inhibition in computing the thermodynamic functions of molecules has been worked out for those cases in which the groups of atoms possessing internal rotation have a symmetry C_n with respect to the axis of rotation, for example, the methyl groups of the ethane molecule, which have a symmetry C_3. The method is due to Pitzer (1), who proceeded from the form of the potential of internal rotation

$$V = \frac{1}{2} V_0 (1 - \cos n\varphi) \qquad (1)$$

Here V_0 is depth of the potential minimum; φ is the angle of internal rotation, and n is the symmetry number. For ethane $V_0 \cong 3300$ cal/mol.

As a result of internal rotations, molecules with groups not having a symmetry C_n may exist in different modifications of varying stability, according to the depth and shape of the potential pits (Fig. 1). These modifications—rotation isomers—were first detected by the Raman spectrum method (2,3). Rotation isomers cannot be distinguished from one another by chemical methods for the time of transition from one form into another is very small—the angular velocity of internal rotation is of the order of 10^{10} sec^{-1}. At the same time the vibration spectra of rotation isomers are variable, because variable are the frequencies, according to the space configuration of the atoms, and the time 10^{-10} sec. is thousandfold or even tens of thousandsfold greater than the time of one vibration.

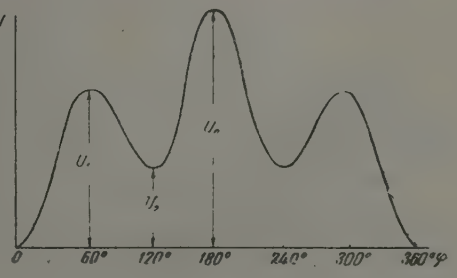

Fig. 1

In computing the thermodynamical functions of rotation isomers account should be taken of the difference in their vibration frequencies. Eliashevich and Stepanov worked out a method of computing and interpreting the vibration spectra of complex molecules, including those possessing rotation isomerism as well (4). This method enables one to compute the vibration entropies of individual rotation isomers, since it gives their separate frequencies. To calculate the entropy corresponding to the internal rotation and torsional vibrations of rotation isomers it is necessary to know the form of the potential. In the case of molecules having

two rotation isomers this potential may be expressed in the same approximation as (1) by the formula

$$V = \frac{1}{6}\{V_0 + 2V_1 + 2V_2 - (2V_0 - 2V_1 + 2V_2)\cos\varphi - \\ -(-2V_0 + 2V_1 + 2V_2)\cos 2\varphi - (V_0 + 2V_1 - 2V_2)\cos 3\varphi\} \quad (2)$$

The values of the potentials V_0, V_1 and V_2 are shown in Fig. 1.

To find the energy levels, and hence the frequencies of the torsional oscillations as well, it is necessary to solve the corresponding problem of quantum mechanics. This leads to Hill's equation, as distinct from the simpler case of C_n symmetry, which leads to the Matier's equation ([5],[6])

$$\frac{d^2\Phi}{d\varphi^2} + \frac{3\pi^2 C}{h^2}\{E - V(\varphi)\}\Phi = 0 \quad (3)$$

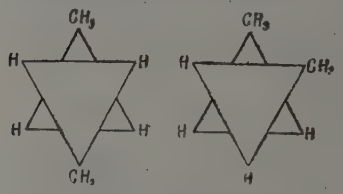

Fig. 2. Left, trans *n*-butane; right, bent *n*-butane

For similarity the following assumptions have been made: 1) The rotator, though asymmetric in the general case, is treated here as symmetrical. 2) The moments of inertia C of the two ends of molecule which revolve about the bond linking them, are supposed to be equal.

To find the frequencies of the torsional oscillations we replace equation (3) by the equation of harmonic oscillators. With this end in viev we will expand the potential (2) in a Taylor series near the points $\varphi = 0$, $2\pi/3$ and $4\pi/3$, and confine ourselves to the second power terms. We get the equations

$$\frac{d^2\Phi}{d\varphi^2} + \frac{4\pi^2 C}{h^2}(E_1 - S\varphi^2)\Phi = 0 \quad (4)$$

$$\frac{d^2\Phi}{d\varphi^2} + \frac{4\pi^2 C}{h^2}(E_2 - Q \pm R\varphi - S'\varphi^2)\Phi = 0 \quad (5)$$

where

$$S = \frac{1}{12}(3V_0 + 24V_1 - 8V_2), \quad S' = \frac{1}{12}(12V_0 + 15V_1 - 23V_2) \quad (6)$$

$$Q = V_2, \quad R = \sqrt{3}\left(V_0 - V_1 - \frac{1}{3}V_2\right)$$

The eigenvalues of the energy are

$$E_1 = h\left(n + \frac{1}{2}\right)\frac{1}{\pi}\sqrt{\frac{S}{C}}, \quad E_2 = V_2 + \frac{R^2}{4S'} + h\left(n + \frac{1}{2}\right)\frac{1}{\pi}\sqrt{\frac{S'}{C}} \quad (7)$$

The approximation we have employed—potential (2) and the conception of harmonic torsional oscillations—is very rough, but may evidently be useful in semi-empirical calculations. We performed such a calculation for *n*-butane molecules.

According to spectroscopic studies, *n*-butane exists in the form of two isomers—a trans-isomer and a bent isomer (gauche) (Fig. 2). Its entropy has been determined by Pitzer([7],[8]) and Aston and Messerly([9]). Pitzer, however, did not take into account the difference in the vibration entropy of rotation isomers, while Aston and Messerly failed to interprete the vibration frequencies properly, and maintained that in an equilibrium mixture there is an excess of the bent isomer. At the same time they assumed the difference in energy of the two isomers $V_2 = 0$

and performed the calculations with the help of the potential of the type (1), putting $V_0 = V_1$. Aston and Messerly found $V_0 = 30\,000$ cal/mol., a value approximately nine times larger than the barrier for ethane. In reality, however, the trans-isomer of *n*-butane is the more stable, which has been proved both experimentally[10] and by computation Eliashevich and Stepanov[11]. The latter found the correct interpretation of the vibration frequencies of *n*-butane. Making use of this interpretation and formulae (8)—(11) (ν_1 corresponds to the trans-isomer, and ν_2 to the bent isomer), we computed the entropy of butane at 272.66 and 298.10 K, the temperatures at which Aston and Messerly determined it experimentally.

The moment of inertia of the methyl group was taken equal to 5.10^{-40} g·cm²; for the ethyl group $C = 60.7 \cdot 10^{-40}$ g·cm². The distance between C—C is 1.53 Å, and between C—H, 1.10 Å; the angles are tetrahedral. The results of a calculation of the entropy not including the entropy of torsional vibrations of the ethyl groups are given in Table 1 (in cal/mol. degree).

Table 1

	272.66°K		298.1°K	
	trans	bent	trans	bent
	Statistical entropy of *n*-butane			
S_{t+r}	60.57	62.00	61.42	62.85
S_v	3.52	4.64	4.29	5.38
S_{ir} ($2CH_3$)	3.84	3.84	4.12	4.12
S'	67.98	70.48	69.83	72.35
	Calorimetrical entropy			
S	72.05±0.2		74.0±0.2	

In Table 1 S_{t+r} is the entropy of the rotation and translation motion of the molecule taking into account the statistical weight of the isomer, which is equal to 1 for the trans-isomer and to 2 for the bent isomer: S_v is the vibrational entropy; S_{ir} ($2CH_3$), the entropy of internal rotation (torsional vibrations) of the two methyl groups at the ends, computed according to Pitzer.

The difference $S - S'$ is due to the torsional vibrations of the ethyl groups and the entropy of displacement. To compute the latter, it is necessary to know the composition of the equilibrium mixture of the isomers of *n*-butane.

There are no such data available, but according to a rough visual evaluation of the relative intensities in the Raman spectra[12], at room temperature the ratio $K = N_{trans} : N_{bent}$ is close to 10. We tested various values of the potentials V_0, V_1 and V_2, and obtained the results given in Table 2, by employing the formulae for the frequencies of the torsional vibrations (V are given in cal/mol.):

$$\nu_1 = \{0.107\,(3V_0 + 24V_1) - 0.85V_2\}^{\frac{1}{2}}\ \text{cm}^{-1} \qquad (8)$$

$$\nu_2 = \{0.107\,(12V_0 + 15V_1) - 2.36V_2\}^{\frac{1}{2}}\ \text{cm}^{-1} \qquad (9)$$

$$\ln K = V_2/RT - \Delta S/R$$

where ΔS is the difference between the total entropies of the two rotation isomers, and

$$S = N_{trans} S_{trans} + N_{bent}\ S_{bent}\ R\,(N_{trans} \ln N_{trans} + N_{bent} \ln N_{bent}) \qquad (11)$$

Table 2

V_0 cal/mol.	V_1 cal/mol.	V_2 cal/mol.	ν_1 cm^{-1}	ν_2 cm^{-1}	$T=272.66°K$		$T=298.1°K$	
					K	S	K	S
3300	**3300**	**2500**	**86**	**60**	**19.3**	**71.73**	**13.3**	**74.37**
3600	3600	2500	93	75	8.4	72.26	6.7	74.47
3600	**3600**	**2200**	**92**	**72**	**12.4**	**71.97**	**9.0**	**74.37**
10000	3300	500	106	130	0.9	73.64	0.8	75.71
»	»	1000	104	126	2.1	73.22	1.8	75.27
»	»	1500	102	121	5.2	73.44	4.2	74.67
10000	**3300**	**2000**	**100**	**116**	**12.9**	**71.88**	**9.4**	**74.11**
10000	5000	500	125	138	0.8	73.80	0.7	75.57
»	»	1000	123	133	1.9	72.92	1.7	75.08
»	»	1500	122	129	4.6	71.84	3.8	74.46
»	»	2000	120	124	11.7	71.62	8.6	73.87
»	10000	1000	167	162	2.6	72.42	1.5	74.70
»	»	2000	164	154	5.9	71.57	7.8	73.42
»	»	3000	161	147	65	70.44	40	72.51
20000	3300	500	120	173	1.0	73.19	0.9	75.26
»	»	1000	119	169	2.4	72.69	2.1	74.79
»	»	1500	117	166	6.2	72.05	5.0	74.20
»	»	2000	115	162	15.4	71.53	11.3	73.71
»	10000	500	178	201	0.8	72.72	0.7	74.87
»	»	1000	177	198	2.0	72.20	1.7	74.37
»	»	1500	176	195	4.9	71.46	3.9	73.70
»	»	2000	174	192	12.1	70.88	9.0	73.12
30000	3300	500	133	207	1.0	72.40	1.0	74.96
»	»	1000	131	204	2.7	72.03	2.3	74.50
»	»	1500	130	201	6.6	71.78	5.3	73.96
»	»	2000	128	198	16.8	71.27	12.4	73.46
Experiment	—	—	—	—	—	**72.05 ±0.2**	**~10**	**74.0 ±0.2**

It is obvious that an exact knowledge of K at different temperatures is necessary for an unambiguous conclusion regarding the values of the potentials.

The values in Table 2 in bold face are in the best agreement with the experimental data. One may in any case feel that the potential $V_0 = 30\,000$, cited by Aston and Messerly, is too high. Nevertheless, it is probable that if not V_1, then V_0 appreciably exceeds the usual value for the C—C bond, close to 3300, for there are steric reasons for this; in the less advantageous cis-configuration of *n*-butane (when $\varphi = \pi$ in Fig. 1), the hydrogens of the two methyl groups may approach one another to a distance of 0.84 A.

For all essential purposes a rigid calculation should take into account such a synchronization of the rotations of the two ethyl groups, at which this situation in general does not occur. Such a calculation, of course, is very complicated.

At the present vibration spectra are the only source of information about the properties of rotation isomers, needed to calculate their thermodynamical properties.

This evidence should not be disregarded by any means. On the contrary, experimental determinations of the relative intensities in the Raman spectra at different temperatures are necessary, for only by this method we obtain the values of the equilibrium constants.

A detailed account of this research will be published elsewhere.

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State Optical Institute.

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30.I.1945.

REFERENCES

[1] K. Pitzer, J. Chem. Phys., **5**, 469 (1937). [2] М. Волькенштейн, Успехим., **13**, 234 (1944). [3] G. Glockler, Rev. Mod. Phys., **15**, 112 (1943). [4] М. Ельяшевич, ЖФХ, (1939—1943). [5] H. Nielsen, Phys. Rev., **40**, 445 (1932). [6] J. Koehler a. D. Dennison, Phys. Rev., **57**, 1006 (1940). [7] K. Pitzer, J. Chem. Phys., **5**, 473 (1937). [8] K. Pitzer, Chem. Rev., **27**, 39 (1940). [9] J. Aston and G. Messerly, J. Am. Chem. Soc., **62**, 1917 (1940). [10] S. Mizusima and al. cited after Chem. Abstr., **64**, 7744 (1940). [11] М. Ельяшевич и Б. Степанов, ЖФХ, **17**, 145 (1943). [12] K. Kohlrausch u. F. Köppl, Z. Phys. Chem. (B), **26**, 209 (1934).

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PHYSICAL CHEMISTRY

MEASUREMENT OF THE IGNITION LIMITS BY THE JET METHOD

By V. KONDRATIEV, Corresponding Member of the Academy

The usual method to determine the ignition limit in a flow of gas is observed by appearance of the flashes. With the temperature of the reaction vessel kept constant, the pressure of the gas is gradually varied, until it reaches the value at which the gas bursts into flame. This value of pressure characterizes the ignition limit at a given temperature.

One of the serious shortcomings of this method is its subjectivity. Investigations of the radiation of hydrogen and carbon monoxide flames show that the output of light is very small in the neighbourhood of the lower ignition limit. Such is apparently the case with other flames, too. That is why visual determination of the flash point is not always reliable, the more so because we often have to work at such temperatures where the furnace heating the reaction vessel and the vessel itself emit a rather intense light. Hence the need for working out objective methods of measuring the ignition limits is obvious.

For reactions accompanied by a change in volume I therefore propose a procedure based on pressure measurements. It was employed by me to determine the lower ignition limit of oxyhydrogen gas in a quartz vessel. The conditions under which flow of gas is created are ordinarily as follows.

The gas (from a gasometer) enters the reaction vessel at a constant rate, practically independent of the pressure in the reaction vessel, which is small in comparison with the pressure in the gasometer. When the evacuation is carried out with oil pumps of the usual type, its rate is proportional to the density of the gas immediately in front of the pump (ρ_k). Denoting the density of the gas in the gasometer and in the reaction vessel by ρ_0 and ρ_1 respectively, the volume of gas leaving the gasometer every second by v_0 (at the pressure and temperature of the gasometer), the volume of gas passing through the reaction vessel per second by v (at the pressure p and the temperature T in the reaction vessel) and the volume evacuated every second by v_k (at the pressure p_k and the temperature T_k), we write the conditions for continuous flow in the following form

$$\rho_0 v_0 = \rho^0 v^0 = \rho_k^0 v_k \qquad (1)$$

$$\rho_0 v_0 = \rho v = \rho_k v_k \qquad (2)$$

Here (1) refers to the case where there is no reaction, and (2), to the case of a reaction in progress.

From equalities (1) and (2) we get

$$\rho^0 \rho v^0 = \rho v \qquad (3)$$

$$\rho_k^0 = \rho_k \qquad (4)$$

Denoting the consumed portion of the hydrogen by γ

$$\gamma = \frac{p_{H_2O}}{p_{H_2} + p_{H_2O}} \tag{5}$$

the ratio of the numbers of moles of oxygen and hydrogen in the mixture by $\alpha/2$

$$\frac{\alpha}{2} = \frac{p^0_{O_2}}{p^0_{H_2}} = \frac{p_{O_2} + \frac{1}{2} p_{H_2O}}{p_{H_2} + p_{H_2O}} \tag{6}$$

and solving jointly the equalities (5) and (5) with the equalities

$$p^0 = p^0_{O_2} + p^0_{H_2} \tag{7}$$

$$p = p_{O_2} + p_{H_2} + p_{H_2O} \tag{8}$$

we get the following expressions for the partial pressures of oxygen, hydrogen and water vapour

$$\left.\begin{aligned} p^0_{O_2} &= \frac{\alpha}{2+\alpha} p^0, \quad p_{O_2} = \frac{\alpha - \gamma}{2+\alpha-\gamma} p, \\ p^0_{H_2} &= \frac{2}{2+\alpha} p^0, \quad p_{H_2} = 2\frac{1-\gamma}{2+\alpha-\gamma} p, \quad p_{H_2O} = \frac{2}{2+\alpha-\gamma} p \end{aligned}\right\} \tag{9}$$

With the help of (9) we find

$$\rho^0 \frac{RT^0}{p^0} = \rho^0_k \frac{RT_k}{p^0_k} = 4\frac{1+8\alpha}{2+\alpha} \tag{10}$$

$$\rho \frac{RT}{p} = \rho_k \frac{RT_k}{p_k} = 4\frac{1+8\alpha}{2+\alpha-\gamma} \tag{11}$$

Substituting (10) and (11) in (3) and (4), we get

$$p = \frac{2+\alpha-\gamma}{2+\alpha} \frac{v^0 T}{vT^0} p^0 \tag{12}$$

and

$$p_k = \frac{2+\alpha-\gamma}{2+\alpha} p^0_k \tag{13}$$

Further, from the Poiseuille formula, assuming the internal friction of the gas to be invariable, we find

$$\frac{v}{v^0}\frac{T}{T^0} = \frac{p^{0^2} - p^{0^2}_k}{p^2 - p^2_k} \frac{p}{p^0} \tag{14}$$

Substituting (14) in (12) and assuming that (in the neighbourhood of the ignition limit) $p^{0^2} - p^2 = 2p^0(p^0 - p)$ and $p^{0^2}_k - p^2_k = 2p^0_k(p^0_k - p_k)$, we get (taking into account (13)) for the relative change in pressure in the reaction vessel the following expression

$$\frac{p^0 - p}{p^0} = \frac{1}{2}\frac{\gamma}{2+\alpha}\left(1 + \frac{p^{0^2}_k}{p^{0^2}}\right) \tag{15}$$

Very frequently a trap cooled by liquid air (or some other cooling mixture) is included in the set-up between the reaction vessel and the pump. In this case an analogous calculation leads to the following expression for the relative change in pressure:

$$\frac{p^0 - p}{p^0} = \frac{1}{2}\frac{\gamma}{2+\alpha}\left(1 + \frac{37+26\alpha}{1+8\alpha}\frac{p^{0^2}_k}{p^{0^2}}\right) \tag{16}$$

From expression (15) and (16) it follows that the maximum change in pressure at a given value of the quantity γ is obtained when $p^0_k = p^0$. In this case we have

$$\delta p = \delta p_k = \frac{p_0 - p}{p_0} = \frac{p^0_k - p_k}{p^0_k} = \frac{\gamma}{2+\alpha} \tag{17}$$

if the reaction products are not frozen out, and

$$\delta p = \delta p_k = \frac{19 + 17\alpha}{(2+\alpha)(1+8\alpha)}\gamma \qquad (18)$$

if they are.

For a stoichiometric mixture ($\alpha = 1$), formula (17) gives $\delta p = \frac{1}{3}$ per cent when $\gamma = 1$ per cent, and formula (18), $\frac{4}{3}$ per cent.

In the case where $p_k^0 \ll p^0$, on the other hand, regardless of whether or not we resort to freezing, we have $\delta p = \frac{1}{2}\frac{\gamma}{2+\alpha}$ (when $\gamma = 1$ per cent, $\delta p = \frac{1}{6}$ per cent). In this case by measuring pressure in front of the pump (p_l) we obtain a very accurate value for the limiting pressure.

The method in carrying out these measurements is as follows. The gas is set flowing through the heated reaction vessel at a certain pressure p^0. Then the temperature of the reaction vessel is gradually raised. As it rises, the pressure remains constant until the temperature of ignition (t_i^0) at this pressure is reached. From then on the pressure falls off (Fig. 1). Thus, by measuring t_i^0, we find the limiting pressure p^0 at this temperature.

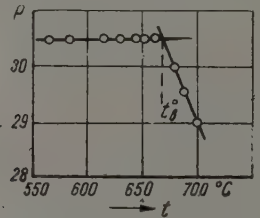

Fig. 1

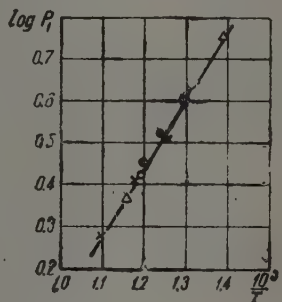

Fig. 2

With the aid of this method we measured the lower limit of ignition of a stoichiometric mixture of hydrogen and oxygen (p_1) in a quartz vessel 2.8 cm in diameter. These measurements were not my primary object, but were merely intended for orientation. No due attention was given, for example, to the accuracy of the measurements. The quantity p_1 was determined with a manometer filled with parafin oil (density 0.88) which indicated the pressure in the reaction vessel. The temperature was measured with a Pt—Ni thermocouple placed in the centre of the reaction vessel.

If we consider the tentative nature of the measurements, the results obtained may be regarded as satisfactory, as the data given in Fig. 2 show. Here the quantity $10^3/T$ has been plotted along the axis of abscissae, and $\lg p_1$, along the axis of ordinates (the values of p_1 have been converted into mm of Hg). The points indicated by crosses, circles, *etc.*, were obtained on different days. The straight line in Fig. 2 fits the formula

$$p_1 = 0.028e^{7700/RT}\ \text{mm Hg}$$

With the special purpose of measuring the lower limit of ignition one may undoubtedly appreciably increase the precision of the method we have employed, especially if measurement of the pressure in front of the pump (p_k) is made the basis of the determination of the flash point. In all probability the upper limit of ignition, too, may be measured by the proposed method.

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CHIMIE DU GLOBE

SUR LA GÉNÈSE DU GERMANIUM DANS LA HOUILLE

Par V. M. RATYNSKI

(Présenté par W. I. Vernadsky, de l'Académie, le 2.XI. 1944)

Le coefficient d'enrichissement du germanium dans le charbon a une valeur maximale, comparée à la valeur d'enrichissement des autres éléments rares, et atteint, comparé au «clark», d'après la terminologie de A. E. Fersman, une valeur de 1600, ce qui attire, naturellement, un grand intérêt sur ce métal. Mais jusqu'à ce jour la question de son origine dans le charbon reste ouverte.

Goldschmidt, qui attira pour la première fois l'attention sur la teneur dans le charbon de toute une série d'éléments chimiques, y compris le germanium, a émis l'opinion dans ses premiers travaux [1-3] que l'accumulation du germanium se produit par adsorption et réduction de ses solutions aqueuses à partir de l'eau circulant à travers les couches de charbon. Dans ses travaux ultérieurs [4] Goldschmidt arrive à la conclusion que l'accumulation du germanium est en relation avec l'activité vitale des plantes ligneuses. D'après l'opinion de Goldschmidt, la concentration du germanium est en rapport avec la quantité de charbons à faible teneur en cendre, comme les claraino-vitrains.

Silbermintz, Roussanov et Kostrykine [5] considèrent que l'origine de la concentration du germanium dans le charbon doit être rapportée aux solutions aqueuses répondant aux minerais effusifs. Se servant des indications de Goldschmidt sur la concentration du germanium dans les minerais de zinc, formés à partir de solutions froides, les auteurs pré-cités expliquent l'accumulation du germanium par la circulation de solutions aqueuses à travers les couches de charbon et par la distribution des minerais sulfureux dans divers endroits de ces couches. Egorov et Kalinin[6] relient l'accumulation du germanium à l'age géologique du charbon considéré.

Ils expliquent la présence du germanium par l'adsorption de ce métal rare par le charbon à partir des roches contenantes et des solutions qui y circulent.

Tous les travaux sur la teneur en éléments rares du charbon n'ont pas tenu compte jusqu'à ce jour de sa structure pétrographique.

Dans nos travaux sur le germanium contenu dans le charbon [7] nous avons étudié un matérial recueilli dans toute une série de bassins houillers: de Kouzbass, d'Oural, de Caucase, etc., et dans la plupart des cas, les espèces étudiées ont été directement prises dans les gisements en tenant compte de la structure pétrographique du charbon. L'étude du charbon du bassin houiller de Khoumarinsk (Caucase), qui est constitué principalement des couches de clarain interposées de feuillets de schiste charbonneux et argileux comprenant des masses de vitrain nous a permis d'aborder la question de la distribution du germanium dans le charbon et d'en arriver ainsi à la question de sa génèse.

Comme nous avons déjà eu l'occasion de l'indiquer [8], on ne peut parler que d'une manière très générale de la concentration du germanium dans les charbons à faible teneur en cendre, comme les claraino-vitrains, si on ne tient pas compte

séparément du clarain et du vitrain. Mais quand on sépare le charbon en ses différents composants, on constate que dans tous les cas la concentration du germanium est liée à une forme pétrographique particulière du charbon, au vitrain.

Tableau 1

Relation entre la teneur en cendre et la concentration du germanium dans les différents composants des divers échantillons de charbon

No. des échantillons	Structure pétrographique	Teneur en cendre en %	Teneur en germanium en % par rapport à la cendre	Relation de la teneur en cendre dans différentes régions d'un même échantillon	Relation inverse de la teneur en germanium dans différentes régions d'un même échantillon
21	Claraino-vitrain avec de couches minces interposées de schiste argileux et des dépôts de sels	18.2	0.008		
21	Claraino-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels. Contient plus de vitrain . . .	13.8	0.026	1.3	3.25
46	Claraino-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	17.3	0.003		
46	Vitrain	3.3	0.113	5.2	37.7
54	Claraino-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	16.1	0.009		
54	Claraino-vitrain aves de minces couches interposées de schiste argileux et des dépôts de sels. Contient plus de vitrain . . .	12.1	0.054	1.3	6.0
55	Claraino-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	34.6	0.003		
55	Vitrain	3.5	0.588	9.9	196.0
160	Clariano-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	9.8	0.004		
160	Vitrain pur	1.4	0.986	7.0	246.5

Quand on compare vitrains et claraino-vitrains pris dans un même échantillon, on voit que la concentration en germanium dans les vitrains est toujours plus marquée que la diminution de la teneur en cendre (tableau 1). Alors que la relation de la teneur en cendre des claraino-vitrains et des vitrains s'exprime par des unités, la relation inverse de leur teneur en germanium s'exprime en dizaines et en centaines (échantillons 46, 55, 160 du tableau 1). La relation de la teneur en germanium dans les claraino-vitrains contenant différentes quantités de vitrains est toujours supérieure à la relation inverse de la teneur en cendre, et est comprise dans les limites d'une dizaine (échantillons 21 et 54 du tableau 1).

On peut supposer que la teneur en germanium dans les claraino-vitrains correspond à la quantité relative de vitrains. La teneur en germanium est toujours plus élevée dans le vitrains même quand on compare des échantillons de vitrain contenant des impuretés et, par suite, à forte teneur en cendre, aux autres composants du charbon donnant aussi peu ou même moins de cendre (tableau 2).

Même dans les variétés les plus pures du claraino-vitrain, ayant une teneur en cendre ne dépassant pas 1.8—2.5 pour cent, la teneur en germanium est d'une dizaine plus faible que les vitrains contenant des impuretés et de teneur en cendre plus élevée.

Tableau 2

Relation entre la teneur en cendre et la quantité de germanium dans les vitrains et les claraino-vitrains

No. des échantillons	Structure pétrographique	Teneur en cendre en %	Teneur en germanium en %, rapportée à la cendre
17	Claraino-vitrains avec de minces couches interposées de schiste argileux et des dépôts de sel	4.3	0.004
24	Vitrain	6.5	0.153
8	Masse de vitrain	3.16	0.168
32	Claréno-vitrène pur	2.5	0.054
45	Claraino-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	3.1	0.011
46	Vitrain	3.3	0.113
55	Vitrain	3.5	0.588
67	Claraino-vitrain pur	1.7	0.051
132	Claairno-vitrain avec de minces couches interposées de schiste argileux et des dépôts de sels	3.7	0.003
190	Vitrain	2.3	0.815

La teneur en germanium dans le vitrain est confirmée également par les résultats de l'analyse des diverses couches de charbon du bassin houiller de Khoumarinsk (8). L'étude des sulfides que l'on rencontre souvent dans le charbon du bassin houiller de Khoumarinsk nous a montré que leur teneur en germanium ne dépasse pas 0.001 pour cent.

Les données que nous avons obtenues sur la distribution du germanium dans les charbons—sur la teneur en germanium du vitrain, nous permettent d'aborder la question de sa génèse. La teneur en germanium du vitrain qui possède la faculté d'absorption la moins élevée, exclut l'absorption de ce métal à partir de ses solutions par un charbon déjà formé. Mais en même temps, au cours de sa formation le vitrain est passé par une étape de hélification la plus complète des végétaux primitifs, et c'est dans cette période qu'a eu lieu vraisemblablement l'absorption du germanium.

L'étude des fusains qui possèdent, comparés à tous les autres composants du charbon, la faculté d'absorption maximale, exclut la possibilité de l'accumulation du germanium par les charbons déjà constitués. Dans aucun cas nous n'avons trouvé d'accumulation du germanium dans les fusains.

Sans s'occuper de la question de l'origine des différents composants des charbons, y compris le vitrain, on peut faire la supposition que les vitrains des différents bassins houillers ont pu prendre naissance de différentes manières, soit par suite d'une différence dans la durée du processus de décomposition d'un même matérial de départ, soit par suite de différents matériaux d'origine.

La condition principale contribuant à la concentration en germanium consiste en un état physico-chimique favorable du milieu aqueux. L'accumulation et la décomposition du matérial végétal, qui a eu lieu dans un milieu aqueux, a conduit à un gonflement des acides humiques dans les premières phases de la formation du charbon. Et plus ce gonflement des acides humiques était intense au cours

des premières phases de la formation du charbon, d'autant plus facilement se faisait la migration vers du germanium, d'autant plus complets étaient les procès de hélification, qui pouvaient conduire à l'absorption du germanium par le gel des acides humiques.

Nous supposons que la teneur en germanium du charbon est due à l'absorption du germanium par le gel des acides humiques à partir des solutions aqueuses à la periode de décomposition des végétaux primitifs. Il ne faut pas exagérer le rôle des solutions enrichies en germanium. Les portions de cette masse, qui sont passées par la phase de hélification la plus poussée, ont donné lieu à la formation du vitrain au cours de la condensation consécutive.

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LITTÉRATURE CITÉE

[1] V. M. Goldschmidt, Nachr. Ges. Wiss. Göttingen (1930). [2] V. M. Goldschmidt u. Cl. Peters, ibid., 141 (1933). [3] V. M. Goldschmidt u. Cl. Peters, ibid., 371 (1933). [4] V. M. Goldschmidt, J. Chem. Soc., 655 (1937). [5] В. А. Зильберминц, А. К. Русанов и В. М. Кострикин, Сб. акад. В. И. Вернадскому к 50-летию научной и перагогич. деятельности, 1936. [6] A. I. Egorov and S. K. Kalinin, C. R. Acad. Sci. URSS, XXVI, No. 9 (1940). [7] В. М. Ратынский, Тр. Биогеохимич. лабор., VIII. [8] V. M. Ratynsky, C. R. Acad. Sci. URSS. XL, No. 5 (1943).

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AGRICULTURAL CHEMISTRY

INFLUENCE OF FEEDING CONDITIONS UPON THE PLANT'S CONTENT OF VARIOUS PHOSPHOROUS COMPOUNDS

By A. V. SOKOLOV

(Communicated by D. N. Prjanischnikow, Member of the Academy, 22.V.1945)

It has been established by means of experiments carried out both with sand and soil cultures that the ratio of the various phosphorous compounds contained in the plant can be varied at will by adding corresponding fertilizers (1). For determining the content of these compounds in plants a special technique has been developed which permits rapidly to evaluate in small weighed portions inorganic phosphates, carbohydrate-phosphates, phosphatids, phytine and nucleoproteid phosphorus (2).

In order to obtain a clearer idea of the phosphorus exchange in plants it is of great importance to make out what is the effect produced by the plant's feeding conditions upon its content of phosphorus compounds; no less important appears the study of the plant's requirement for phosphorus, and the development of procedures likely to increase the quality of the yield. The evaluation of the amount of inorganic phosphates contained in young plants is necessary in order to determine the plant's requirement of phosphatic fertilizers (3). The determination of inorganic phosphorus is a much more rapid and easier procedure than that of total phosphorus; its results, however, do not always furnish reliable information as to the real supply of the plant with phosphorus. An increase in the dose of phosphate fertilizers is not necessarily followed by a corresponding increase in the percentage of inorganic phosphorus in the plant. A liberal phosphatic diet leads to an intensified growth of the plant, as a result of which the phosphates supplied to the plant are more intensively consumed for the formation of organic phosphorous compounds.

Culture experiments carried out with maize grown on common black soil have shown that an increase in the dose of phosphate from 0.05 g of P_2O_5 to 0.20 g introduced per 1 kg of soil, failed to produce any increase in the percentage content of inorganic phosphates in the experimental plants. In the experiment described below the increase of the dose of prosphates brought about a vigorous growth of the young maize plantlets (Table 1).

Table 1

Content of Inorganic Phosphates in Young Maize Plants

Dose of P_2O_5, g/kg	Yield of dry matter, g	Inorganic phosphates P_2O_5, %
0.00	12.1	0.186
0.05	16.8	0.205
0.10	24.1	0.197
0.20	32.6	0.201

A more liberal phosphate diet supplied to flax and barley plants during the period of intense growth also failed to augment the quantity of inorganic phosphates contained in these plants (1), although their rate of growth was increased considerably. In another experiment, carried out with maize grown on podzolized soil, heavy doses of P_2O_5 were administered, amounting to 0.20 g of the fertilizer per 1 kg of soil. Here again an important increase in the yield of young plants was obtained, but there was no rise in the ratio of inorganic phosphates. Further increase in the dose of P_2O_5 ceased to produce any important increment in the yield, but the content of inorganic phosphates in the plants was now enhanced considerably. The percentage of organic phosphates was found to vary but little with the increase in the dose of phosphatic fertilizers supplied; the ratio of organic phosphates to inorganic ones was thus found to shift to the advantage of the latter (Table 2).

Table 2

Ratio of Inorganic to Organic Phosphates in Maize

Dose of P_2O_5, g/kg	Yield of dry matter, g	Inorganic phosphates P_2O_5, %	Organic phosphates P_2O_5, %	Organic phosphates in % to inorganic ones
0.0	6.3	0.180	0.185	103
0.2	21.7	0.173	0.196	113
0.4	28.0	0.317	0.198	62
0.8	27.9	0.566	0.217	38
1.6	28.7	0.796	0.222	28

From this table it may be seen that under conditions of an excess of phosphatic fertilizers the content of inorganic phosphates becomes a reliable index of the plant's provision in phosphorus. Under conditions of rapid growth the scarcity of inorganic phosphates should not be interpreted to mean that the plant suffers from phosphorus deficiency.

The most important phosphorus compounds contained in seeds are phosphatids, phytine and nucleoproteids. By altering the conditions of phosphorus nutrition it is possible materially to affect the plant's content in nucleoproteid phosphorus, and especially in phytine. In our culture experiment with oats, carried out on sand cultures, at a stable pH =6.0—6.5, in addition to one of the two forms of nitrogen, the nitrate or the ammoniacal, two doses of P_2O_5 were introduced, one of 200 mg and another of 40 mg per pot. With either form of nitrogen the P_2O_5 deficiency (40 mg) resulted in an abrupt drop in the content of phytine in the seeds, in an essential reduction in the nucleoproteid phosphorus, while the amount of phosphatids in the grain remained but slightly affected (Table 3).

Table 3

Effect of Fertilizers upon the Content of Phosphatids, Phytine and Nucleoproteids in the Grain of Oats

Nitrogenous fertilizers	Dose of P_2O_5 in g per pot	P_2O_5 in %		
		Phosphatids	Phytine	Nucleoproteids
$Ca(NO_3)_2$	0.20	0.095	0.310	0.343
»	0.04	0.112	0.085	0.180
$(NH_4)_2SO_4$	0.20	0.100	0.226	0.417
»	0.04	0.110	0.090	0.290

The change in the plant's provision with phosphates has thus been found materially to affect the content of nucleoproteids and particularly of phytine in the grain, whereas it failed to affect the content of phosphatids. But the content of phosphatids in the young oat plants was essentially affected by the change in the conditions of their phosphate nutrition. The same culture experiment involved the determination of phosphatids in 30-day-old oat-plantlets (Table 4); from the results of these determinations it may be seen that the content of phosphatids in plants may vary greatly according to food conditions.

Table 4

Effect of Fertilizers upon the Content of Phosphatids and Inorganic Phosphates in Plants

Nitrogenous fertilizers	Dose of P_2O_5 in g per pot	Phosphatids in % of P_2O_5	Inorganic phosphates in % of P_2O_5
$Ca(NO_3)_2$	0.20	0.090	0.531
»	0.04	0.048	0.110
$(NH_4)_2SO_4$	0.20	0.107	0.909
»	0.04	0.054	0.109

The content of the different forms of phosphorus compounds in plants is not affected in a similar way by ammonia or nitrate diet [5]. An adequate supply of phosphates is necessary for the reduction of nitrates in the plant. Thus, Kraybill [4], Eckerson [5] and Turchin [6] have shown that in the case of nitrate diet phosphorus deficiency leads to the accumulation of nitrates and carbohydrates, on the one hand, and to nitrogen starvation of the plant, on the other. It has, therefore, been suggested by Eckerson and Turchin that there is a definite relation between the amount of phosphates present in the plant and the intensity of the reduction processes that take place in the latter. But at the time being we are in position to give this phenomenon a different interpretation. During the last few years much has been done to shed light upon the rôle of phosphorus compounds in the carbohydrate metabolism and the biological redox processes. The reactions of phosphorylation have been given thermodynamic interpretation by Kulkar [7] and Lipman[8], independently of each other. It has been found that compounds with a phosphorus bond high in energy play an important part in biological processes of phosphorylation. The reactions of dehydration of aldehydes, keto-acids and unsaturated acids are accompanied by the formation of carboxyl phosphates and ionol-phosphates which can transmit their phosphorus group ($\sim$pH) to adenyl acid, transforming it into adenosindi- and triphosphoric acids. Thus, phosphorus is necessary for the oxidation of carbohydrates, as a result of which reduction of nitrates by the hydrogen freed ($2H^{\cdot} + 2e$ or $3H^{\cdot} + 2e$) takes place as a concomitant reaction, with a simultaneous formation of organic acids and CO_2 from the hydrocarbons. The interpretation here suggested is in good agreement with the data on the influence of nitrate diet supplied to plants upon the formation of organic acids[9]. At the present time we are engaged in a study of the rôle of P_2O_5 in the nitrate nutrition of plants, which we suppose to clarify with the aid of an analysis of the content of the various forms of phosphorus compounds in our experimental plants.

Conclusions. 1. The conditions of phosphorus nutrition may materially affect the rate of accumulation of the various forms of phosphorus compounds in the plant.

2. The content of inorganic phosphates in the plant depends not merely upon

their content in the nutrient medium, but also upon the rate of their uptake by the plant at the different periods of growth.

3. The degree to which the plant is supplied with phosphorus may essentially affect the content of phytine and nucleoproteids in the grain, as well as the content of phosphatids in young plantlets.

4. The participation of phosphates in the reduction of nitrates in the plant can be brought about by their participation in the oxidation of carbohydrates, owing to which the carbohydrates are oxidized into organic acids and CO_2, while nitrates are reduced by the hydrogen split-off the carbohydrates.

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REFERENCES

[1] А. В. Соколов, Химизация соц. земледелия, № 8, 22 (1939). [2] А. В. Соколов, там же, № 10,36 (1940). [3] В. Н. Панфилов и А. В. Соколов, Диагностика питания растений, 1944. [4] H. K Kraybill, Ind. and Eng. Chem., 22, 275 (1930). [5] Sophia H. Eckerson, Contr. Boyce Thompson Institute, 2, 197 (1931). [6] Ф. В. Турчин, О природе действия удобрений, 1936. [7] H. M. Kulkar, Chem. Rev., 23, 71 (1941). [8] F. L. Lipman, Advances in enzymology and related subjects. 1, 99 (1941). [9] А. В. Владимиров, Химизация соц. земледелия, № 8, 35 (1939),

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GEOLOGY

A SCHEME OF DIVISION OF THE CRETACEOUS DEPOSITS OF FERGHANA

By O. S. VIALOV

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The present paper contains a very concise characteristic and subdivision of the main types of the Cretaceous sections of Ferghana and their correlations.

Two main types will be discussed: the Isphara type characteristic of south Ferghana, and the Kizyl-Yar type embracing the sections of north-east Fergahna. It is to these sections that all the other Ferghana sections are to be correlated.

The Isphara type

A. Guzan series [members I—III(²)], unconformably overlying the Rhaetic-Jurassic beds and having at its base basal conglomerates (10 m), consists of red marlaceous clays with a large quantity of white marlaceous concretions, sometimes overcramming the rock, and greenish-grey cross-bedded sandstones, forming along with bands separate thick beds (up to 8 m). Thickness, about 50 m.

B. Lyakan series (member IV). A series persistent and sharply distinguishable in the relief, consisting of white and pink dolomitized limestones and dolomites with a cavernous weathering surface. Thickness, 25 m.

C. Kizyl-Pilyal series (members V—VII). A series of red marlaceous clays with a varying quantity of marl concretions. It is subdivided into two parts by a member of carbonate rocks, consisting in the lower part of a dense, almost black limestone (4 m), overlain by ligth-coloured whitish porous marls (10 m). Thickness, 52 m.

D. Kansk series (member VIII). Conglomerates alternating with red clays (13 m) and overlain by a series of red clays (10 m). Thickness, 23 m.

E. *Exogyra* series (member IX—X) begins with a bed (1.5m) of conglomerate passing along the strike into sandstone and containing a comparatively small quantity of oysters: *Exogyra* cf. *turkestanensis* Born. and *Liostrea* sp. Higher up follows a member consisting of cloddy marls with numerous bands of red and green marly clays; there are also bands of sandstones passing into gravel bands and rarely, into fine pebbly conglomerates. Thickness, 15 m.

The *Exogyra* series in the Isphara section is by no means typically represented. Strictly speaking, it is only a stratigraphic analogue of the true *Exogyra* series overfilled with oysters, which is developed in eastern Ferghana.

F. Yalovach series (member XI). A characteristic series of red or pink banded sands, wery slightly cemented together. A discontinuous horizon (10—15 cm) was discovered within these sands near the roof in 1943, containing a very fragile fauna of pelecypods and numerous fragments of bones of dinosaurs and shields of turtles and crocodiles. Thickness, 25 m.

G. Variegated series (members XII—XIV). It is a series of rather closely alternating sands, sandstones and clays red and greenish, yellow

Subdivision of Sections of the Isphara Type

Vialov, 1943			Kalitzky, 1915			Arkhangelsky, 1916
Suzak stage	Gypsum with clay bands	SzK		Gypsiferous formation		Sn
Bukhara stage	4. Limestone with a typical fauna 3. Fish horizon of shales and sandstones 2. Sulyukta sandstones 1. Green clays	Bch	k			
Danian stage	Goznau gypsum	H	i			
Senonian	Variegated series	G_3 G_2 G_1	— h — — g —			
Cenomanian-Turonian	Yalovach sand series	F_0		Sands		T?
	Exogyra series	E	— f — — e —	Cretaceous deposits	Series of clays with shell-rock bands	
	Kansk conglomerate series	D	d		Red clays and conglomerates	Cm
Lower Cretaceous?	Red-coloured Kizyl-Pilyal series	C	— c —			
	Lyakan limestones	B	b		Dark-red series of clays with beds of limestones b and c	Cr_1
	Red-coloured Guzan series	A	— a —			

and grey. In the lower part distinctly predominate sandy bands, and only in the upper part clays appear in an appreciable quantity. In the uppermost portion the sandstones occupy a subordinate place, but among the greenish, reddish and almost white clays and marls (and clayey dolomites) there occur numerous bands of gypsum and a few thin horizons of white porous dolomitized limestones, sometimes overcrammed with a uniform inderterminable fauna of pelecypods and gastropods.

If the exposures of the series here described be viewed at a distance, one is struck in the first place by its fine and variegated banding. In the lower part red shades are distinctly predominant (I member), and upward they gradually assume, first, a general pink, and afterwards, a greenish and whitish hue. Total thickness of the series, 67 m.

Subdivision of Sections of the Kizyl-Yar Type

	Vialov, 1943		Bezobrazova, 1929	Pavé, 1933	Nalivkin, 1926	Arkhangelsky, 1916	Bronnikov, Weber and Faas, 1910	
Suzak stage	Green clays. Limestone with oysters. Green clays	SzK	Lower Eocene	Danian stage	Suzak horizon	SzK	1 2	First gypsiferous series
Bukhara stage	Gypsum beds with bands of limestones	Bch	Palaeocene	Danian stage	Senonian gypsiferous stage	Sn. s	3	First gypsiferous series
Danian stage	Red clays overlying the radiolite horizon	H	Upper Senonian (Danian stage)	Upper Senonian			4	
Maestrichtian	Radiolite horizon	G_3	Santonian	Santonian	Radiolite horizon		5 6	Radiolite horizon
Senonian	Tekebel red sands and clays	G_2	Turonian		Brown clays and sandstones with bands of white limestones	Snt	7 8 9 10	Clayey-calcareous series
	Aga-Aral gypsiferous horizon	G_1			Turonian gypsiferous series	T	11 12 13	Second gypsiferous series
Cenomanian-Turonian	Yalovach series of red sands	F			Variegated areno-argillaceous series		14 15 16	Sandstones above the oyster beds
	Exogyra series	E	Cm		*Exogyra* series	Cm	17* 18 19 20 21 22	Oyster beds
Lower Cretaceous?	Changet series	D | A	Cr_1		Lower Cretaceous red sandstones and conglomerates	Cr_1		

* From the description of the talus of layer 17 it is difficult to decide whether this part of the section enters the Yalovach or the *Exogyra* series.

Comparison of Sections

	Isphara type				Kizyl-Yar type	
Bukhara stage	Bch		4. Limestones with a typical fauna 3. Fish horizon of shales and sandstones 2. Sulyukta sandstones 1. Green clays	Bch	Formation of white gypsum with bands of limestones with a typical fauna	
Danian stage	H		Goznau gypsum	H	Series of red clays overlying the radiolite horizon	
Senonian	G_3	Variegated series	Gypsiferous member	G_3	Variegated series	Radiolite horizon
	G_2		Variegated sands and clays	G_2		Tekebel red sands and clays
	G_1		Red sands and clays	G_1		Aga-Aral gypsiferous horizon
Cenomanian-Turonian	F		Yalovach sand series	F	Yalovach sand series	
	E		*Exogyra* series	E	*Exogyra* series	
	D		Kansk series of conglomerates	A—D	Changet red-coloured series	
Lower Cretaceous?	C		Kizyl-Pilyal red-coloured series			
	B		Lyakan limestone			
	A		Guzan red-coloured series			

H. Goznau gypsum series (member XV) consists in the lower part of gypsum beds (1—1.5 m) divided by small riders of closely alternating layers of gypsum and greenish-grey, olive and red clays. Higher up follows a continuous series of white finely crystalline gypsum. Thickness, 26 m. Along the Goznau gypsum passes the boundary of the Cretaceous and the Palaeogene (3).

The Kizyl-Yar type

A—D. Changet series. It is a very thick and highly uniform series composed mainly of red-coloured rocks, clays and sandstones, among which now and then appear bands of green clays, as well as of green and grey sandstones, not infrequently passing into conglomerates.

As distinguished from Isphara, here Lyakan limestones are absent (as well as limestone bands in general), and the Changet series undergoes no further sub-

division. Moreover, the Kansk series (D), which is fairly well distinguishable in south and partly in south-east Ferghana, cannot be separated here from the Changet series. Thickness, up to 1200 m.

E. *Exogyra* series. In the lower portion there is an alternation of continuous oyster beds (up to 3 m thick) with thin layers of green clays. Subordinate to these beds are dense intercalations (up to 85 cm) of light-grey and greenish sandstones and marls, occasionally passing into limestones, containing a large amount of similar oysters. In the upper portion of the series green clays are predominant. From the present author's collections B. A. Borneman has determined: *Liostrea* cf. *delettrei* Coq., *L. oxiana* Rom., *L.* cf. *thevestensis* Coq., *Gryphaea biauriculata* Lam., *Exogyra columba* Lam., *Ex.* cf. *africana* Lam., *Ex. olisiponensis* Sharpe, *Ex. turkestanensis* Bornem. Among the other pelecypods, L. V. Romanovskaya has determined: *Plicatula* cf. *auressensis* Coq., *Modiola vialovi* sp. n., and *Callistina* cf. *plana* Sow. To this list may be added the following species, previously determined by A. D. Arkhangelsky(1): *Trigonia turkestanensis* Arkh., *Plicatula* cf. *batnensis* Coq. and *Placenticeras stantoni* Hyat. Thickness, 60 m.

As compared to the *Exogyra* series from Isphara, the thickness is considerably increased here, and the oyster fauna becomes decidedly more abundant and diversified. As distinguished from the sections of south-east Ferghana, absent from Kizyl-Yar are the characteristic beds of limestone, traceable there in the roof and in the base of the series.

F. Yalovach series. This is a red-coloured sandy series which may be subdivided into two parts (in ascending order):

1. A member consisting of green and red spotted clays and sands; in the roof occurs a very dense calcareous dark-red sandstone (0.35 m) with numerous sections of shells. This member is provisionally referred to the Yalovach series. Thickness, 25 m.

2. A member consisting of coarse-grained red, pink and light-grey sands, with bands of red spotted clays and thin crests of sandstones of a denser consistency. Thickness, 195 m.

It is of interest to recall here the finding of a fragment of bone in the upper part of this series, reported to by the authors of the «Andizhan earthquake».

G. Variegated series. It is easily subdivided into three series, which distinguishes it from the sections of south-east Ferghana. As already shown, three series are also distinguished in the Isphara section, but on the basis of quite different features, for there they have a different composition. The presence of the radiolite horizon is one of the most characteristic distinctive features of Kizyl-Yar.

G_1. Aga-Aral gypsiferous horizon. Green and red clays with some marls, grey sandstones and gypsum. Upon a more detailed examination three members are distinguished:

1. Green and grey clays, in the lower part with sparse bands of sandstones and marls, in the middle portion alternating with marls, and in the upper part, with sandstones. The clays contain a vast amount of very fragile impressions of pelecypods, especially in the upper part of the member. Thickness, 20 m.

2. Red clays with gypsum beds (only in one case up to 1.30 m) and numerous thin bands of light-grey gypsified sandstones (of the type of the Fontainebleau sandstones). Thickness, 17 m.

3. Green clays with gypsum beds (up to 70 cm) and with numerous thin bands of sandstones. In the roof lies the thickest bed of gypsum (1.5 m). Thickness, 15 m.

G_2. Tekebel horizon of red sands and clays. An argillo-arenaceous horizon consisting of two parts:

1. Red argillaceous sandstones and arenaceous clays with a few bands of green clays and light-coloured somewhat denser sandstones. Thickness, 35 m.

2. Light-coloured sands, white, yellowish and greenish, with sparse, thin bands of red sandy clays. Thickness, 11 m.

G_3. Radiolite horizon. The radiolite horizon consists of several beds of limestones separated by layers of red clays. The upper horizon of lime-

stones, which is the thickest one (9 m), holds remains of radiolites. In general, the quantity and thickness of the limestones varies, and the limestone in the lower horizon sometimes passes into sandstone.

To the list of fauna given by Arkhangelsky([1]), Vialov has added on the basis of the fauna collected by Bezobrazova: *Procassidulus* aff. *umbonatus* Woods, and *Procassidulus* (*Australanthus*) sp. indet. ([4]).

H. Series overlying the radiolite horizon. It is a series of dark-red clays, brown-red and brownish, sometimes arenaceous, highly gypsiferous, with a large quantity of rounded concretions and veins of gypsum, with sparse thin intercalations and lenses of grey, greenish sandstones. In the topmost portion appear pink sands and pink clays (9.40). The total thickness is about 100 m. Higher up follows gypsum with bands of limestones, referred to the Bukhara stage of the Palaeogene.

The presence of the red-coloured argillaceous series overlying the radiolite horizon is one of the principal points of distinction of the Kizyl-Yar section from those of south Ferghana, where the stratigraphic analogues of this series are represented by a member composed of continuous gypsum (the Goznau gypsum series).

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REFERENCES

[1] А. Д. Архангельский, Тр. Геол. ком., в. 151 (1916). [2] О. С. Вялов, Мат. ТПВ, в. 47 (1936). [3] O. S. Vialov, C. R. Acad. Sci. URSS, XLII, No. 2 (1944). [4] О. С. Вялов, Изв. Гл. геол.-разв. упр., L, в. 4 (1931). [5] К. П. Калицкий, Тр. Геол. ком., в. 133 (1915). [6] Ф. Н. Чернышев, М. М. Бронников, В. Н. Вебер и А. В. Фаас, Тр. Геол. ком., в. 54 (1910).

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GEOLOGY

THE BASHKIRIAN STAGE OF THE MIDDLE CARBONIFEROUS AND ITS SUBDIVISION

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The «Bashkirian beds» with a Middle Carboniferous fauna older than that of the Moscovian stage were separated out by Semikhatova in the Urals in 1934 [8]; these beds are absent from the near-Moscow region. Since that time geologists have expressed different views concerning the Bashkirian beds: some of them regarded them as a large stratigraphic complex of the type of the C_2 stage [8,13]; others conventionally considered them as equivalent to the Namurian stage[4]; by others they were viewed as a facial type of the lower beds of the Moscovian stage [1]; others still assigned to them a stratigraphic value of a restricted horizon arbitrarily referred to the lower portion of the Middle Carboniferous $=C_2^\circ$ [3]. Thus, there was no clear idea as to the stratigraphic scope of the Bashkirian beds. Finally, Stepanov [11], in a section of the Carboniferous on the Zigan River (Bashkiria), separated out the «Bashkirian series», C_2^1, as corresponding to the lower part of C_2, absent from the Moscow basin, and to the upper part of the Namurian stage.

The material accumulated by now on the fauna, stratigraphic significance and distribution of the Bashkirian beds in the USSR makes one recognize them as an independent stratigraphic unit of the same order as the Moscovian stage. This material, part of which is presented below, justifies one to separate out the so-called Bashkirian beds as an independent Bashkirian stage of the Middle Carboniferous (C_2^1) as distinguished from the Moscovian stage (C_2^2).

The standard section permits of a distinct subdivision of the Bashkirian stage, which is composed there of a series of carbonate rocks, into three horizons (in the ascending order). It is this section that served once to evaluate for the first time the age of the Bashkirian beds as being significantly older than the Moscovian stage [8,13]. The latter has been studied in detail by the present author [12,13], while its choristites have been described in a monograph by Semikhatova [10].

1. Lower Bashkirian horizon C_2^{1a}, about 20—30 m thick, with *Thysanophyllum tschernowi* Soschk., *Spirifer* sp. ex gr. *striatus* Mart., *Linoproductus* aff. *ovatus* Hall., *Martinia robusta* Semich., *Derbyia* cf. *grandis* Waag., *Staffella antiqua* Dutk., *St.* ex gr. *bradyi* Möll., *St.* ex gr. *struvei* Möll., *St. nana* Raus., *St. pseudostruvei* Raus. et Bel., *Archaediscus baschkiricus* Krest. et Teod., *Bradyina* cf. *cribrostomata* Raus. et Reith.; in the upper part of the horizon or at its boundary with the overlying one on the Basa River (a left-hand tributary of the Great Inzer River) *Choristites lebedevi* Fred. has been encountered.

2. Middle Bashkirian horizon (C^{1b}), about 50 m thick, with *Choristites bisulcatiformis* Semich., *Ch. bisulcatiformis* Semich. var. *anomale* Semich., *Ch. yanghukouensis* Chao var. *crassicostata* Semich., *Ch. robustus* Semich., *Ch. urtaayryensis* Semich., *Ch. yakhiaensis* Semich., *Ch. lebedevi* Fred., *Ch. lebedevi* Fred.

var. *laticostata* Semich., *Ch. uralicus* Leb. var. *asiatica* Fred., *Ch.* cf. *holtedahli* Fred., *Spirifer* aff. *bisulcatus* Sow., *Linoproductus* aff. *ovatus* Hall., *L. cora* d'Orb. var., *Derbyia* aff. *crassa* Meek and Hayd., *Staffella antiqua* Dutk., *St. confusa* Lee et Chen., *St.* ex gr. *bradyi* Möll., *St.* ex gr. *struvei* Möll., *St. nana* Raus., *St. pseudostruvei* Raus. et Bel., *St. angulata* Col. (upper half of the horizon), *Bradyina* cf. *cribrostomata* Raus. et Reitl., *Archaediscus baschkiricus* Krest. et Teod.

3. Upper Bashkirian horizon (C_2^{1c}), about 30 m thick, with *Choristites yanghukouensis* Chao var. *baschkirica* Semich., *Ch.* cf. *fascilatus* Semich., *Ch.* cf. *holtedahli* Fred., finely costate *Choristites* sp., *Spirifer* aff. *bisulcatus* Sow., *Echinoconchus punctatus* Mart., *Eomarginifera* cf. *lobata* Sow., *E.* cf. *sapojnikovi* Semich., *Marginifera* cf. *schartimiensis* Jan, *M. loczii* Chao., *Profusulinella* ex gr. *praecursor* Deprat., *Staffella angulata* Col., *St.* ex gr. *antiqua* Dutk., *St.* ex gr. *bradyi* Möll., *St.* ex gr. *struvei* Möll., *Archaediscus baschkiricus* Krest. et Teod.

All the three horizons of the Bashkirian stage (C_2^{1a}, C_2^{1b}, C_2^{1c}) contain a Middle Carboniferous fauna older than that of the Moscovian stage. The most typical Bashkirian complex of fauna (*Choristites* ex gr. *bisulcatiformis* Semich., abundant *Staffella antiqua* Dutk. and *Archaediscus baschkiricus* Krest. et Teod.) is peculiar to the Middle Bashkirian beds composing the middle, thickest, part of the section of the stage. Besides the presence of typical Bashkirian forms, the upper Bashkirian beds are distinguished by the appearance of the first *Profusulinella* and the first representatives of finely costate *Choristites*. The lower member of the section described, corresponding to the beds with *Staffella antiqua* Dutk. (conventional «Upper Namurian» of certain geologists), contains a number of representatives of the Middle Carboniferous fauna (in particular, the complex of *Foraminifera* in this member is identical with Middle Bashkirian one. Consequently, it should be referred to the lower part of the lower or Bashkirian stage of the Middle Carboniferous. In the section of the Zilim the Upper Bashkirian horizon is overlain by carbonate rocks with a fauna of the Moscovian stage ([8],[13]).

In the Shikhany—Ishimbay region of the near-Ural area the same three horizons of the Bashkirian stage have been established; they are composed of carbonate rocks and are characterized in general by the same faunistic complex (according to data by V. A. Balaev, D. F. Shamov and A. V. Pakhomova):

a) The lower horizon (C_2^{1a}), 35 m thick, with *Chonetes carbonifera* Keys., *Linoproductus cora* d'Orb., *L.* aff. *ovatus* Hall., *Martinia baschkirica* Hall., *Derbyia* cf. *regularis* Waag., *Archaediscus baschkiricus* Krest. et Teod.

b) The middle horizon (C_2^{1b}), 135 m thick, with *Choristites* cf. *bisulcatiformis* Semich., *Ch. ishimbaevi* Bal., *Ch., preuralica* Bal., *Ch. laticostatus* Bal., *Ch. ornatus* Bal., *Chonetes carbonifera* Keys., *Linoproductus cora* d'Orb., *Echinoconchus* cf. *punctatus* Mart., *Derbyia regularis* Waag., *Staffella* ex gr. *antiqua* Dutk., *St. antiqua* Dutk. (*St. prosphaeroidea* Raus.), *Archaediscus* ex gr. *baschkiricus* Krest. et Teod.

c) The Upper horizon (C_2^{1c}), 50 m thick, with *Choristites* sp. nov. ex gr. *bisulcatiformis* Semich., *Ch. minimus* Bal. (the first representative of finely costate *Choristites*), *Chonetes carbonifera* Keys., *Linoproductus cora* d'Orb., *Krotovia* cf. *rarituberculata* Ivan., *Marginifera baschirica* Ivan., *Profusulinella parva* Lee et Chen., *P. prisca* Deprat.

Like in the Urals, in the Ishimbay region the most typical Bashkirian faunistic complex corresponds to the middle, thicker portion of the section of the Bashkirian stage, while in its upper portion also appear the first *Profusulinella* and finely costate *Choristites* (see also([14])). The lower horizon of the section also refers to the base of the Bashkirian stage from the appearance of a number of representatives of the Middle Carboniferous fauna. The Upper Bashkirian horizon in the Ishimbay region is also overlain by carbonate rocks with a fauna of the Moscovian stage.

The three horizons of the Bashkirian stage separated out in the Urals and in the near-Ural area may be traced in the sections of the Carboniferous of the eastern half of the Russian platform. An analysis of the latter sections has shown:

a) at the boundary of the Lower and Middle Carboniferous the sea of the Russian platform receded East beyond the line Syzran—Buldyr—Krasnokamsk; b) three progressive movements from E to W (and then regressive ones) of the Bashkirian sea had taken place, to which correspond the deposits of the three horizons of the Bashkirian stage separated out by the author; c) the regression at the boundary of the Bashkirian and Moscovian ages had been a considerable one.

The greatest thicknesses (up to 220 m) and a continuous section of the Bashkirian stage are observed in the Sterlitamak-Isihimbay region; and a somewhat lesser thicknesses (110 m), still farther East, on the western slope of the Urals (the Zilim River, *etc.*). In both regions the accumulation of marine carbonate sediments proceeded throughout the Bashkirian age without interruption. On the contrary, in the West of the territory considered the thicknesses of C_2^1 are small. Shiftings of the Bashkirian sea which had taken place here could have been accompanied by an emergence of some regions. Finally, still further West, beyond the trans-Kama Tataria, stretched a land whose eastern border only was occasionally submerged by the Bashkirian sea.

Deposits of the Bashkirian stage are known not only from the Urals, the near-Ural area and the eastern half of the Russian platform, but also from the Donetz basin ([3],[7]), Ferghana([2]) and some other regions. In the section of the Carboniferous of the Donetz basin, where the faunistic and floristic evidence are open to comparison, it may be seen that the two upper *Goniatites* zones of the Namurian of Western Europe (with *Reticuloceras superbilingue* Bisat. and with *Gastrioceras rurae* Schmidt) are to be referred to the foundation of the Westphalian deposits of Europe, because of the re-appearance in these *Goniatites* zones of the fauna of brachiopods (the presence of *Choristites* of a Bashkirian habit), *Foraminifera*, and a wide development of a typical Westphalian floristic complex([7]). These two *Goniatites* zones correspond to the lower part of the Bashkirian stage.

Thus, the material here discussed gives us right to consider the so-called «Bashkirian beds» as an independent—Bashkirian—stage of the Middle Carboniferous, and to distinguish within this stage three stratigraphic horizons.

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REFERENCES

[1] V. A. Balaev, Bull. Acad. Sci. URSS, sér. géol., No. 5 (1940). [2] V. A. Vakhrameev, ibid., No. 2 (1938). [3] N. P. Malakhova, ibid., No. 4—5 (1941). [4] Д. Наливкин, Путеводитель XVII сессии Международн. геолог. конгресса, 1937. [5] Д. В. Наливкин. Геолог. история Урала, изд. АН СССР, Свердловск, 1943. [6] D.M. Rauser-Chernoussova, Bull. Acad. Sci. URSS, sér. géol., No. 2 (1943). [7] А. П. Ротай, Труды XVII сессии Международн. геолог. конгресса 1937, 1, М., 1939. [8] С. В. Семихатова, Проблемы сов. геологии, IV, № 8 (1934). [9] С. В. Семихатова, Бюлл. Моск. об-ва испыт. прир., отд. геол., XIV, № 3 (1936). [10] С. В. Семихатова, Труды Палеонт. ин-та АН СССР, XII, в. 4 (1941). [11] Д. Л. Степанов и С. Н. Гусева, Тр. НГРИ, сер. А, в. 115 (1939). [12] Г. И. Теодорович, Бюлл. Моск. об-ва испыт. прир., XIII, № 1 (1935). [13] Г. И. Теодорович, Проблемы сов. геологии, VI, № 7 (1936). [14] И. В. Хворова, Бюлл. Моск об-ва испыт прир., отд. геол., XV, 6 (1937).

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MINERALOGY

ON THE SECONDARY ALTERATION OF PYROCHLORE

By G. P. BARSANOV

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The mineralogical literature is lacking any data as to the processes of secondary alteration in minerals of the group of titano-niobates, the composition of their alteration products, and, consequently, the processes responsible for them. However, the existence of secondary alteration on a wide genetic scale of a pegmatite process is beyond any doubt in the case of minerals of the pyrochlore group. In the pegmatites of the Ilmen Mts., which are the best known deposits of this mineral in the USSR, the octahedral faces of crystals of pyrochlore are almost always covered with an opaque crust, creamy or brownish-yellow in colour, while crystals with lustrous faces are exceedingly rare. Although the presence of these crusts was recorded nearly by all mineralogists who studied the pyrochlores of the Ilmen Mts., the composition of the secondary mineral, its properties and conditions of formation have not been studied by any of them. Its study offers a number of essential difficulties, indeed, for the thickness of the secondary crust does hardly attain 0.1 mm.

The author succeeded in finding among the materials of the Ilmen National Park a specimen with crystals of pyrochlore, which had been subject to rather intense processes of secondary alteration; on the largest crystal (about 4 mm along the edge) a complex reticulate corrosion texture had formed, while the smaller crystals (about 1 mm) had entirely passed into secondary products. The specimen is a piece of coarse-grained feldspar-biotitic pegmatite, probably originating from the group of Cheremshan mines of the syenite zone (No. 107 or 108). In the hand specimen a miarolitic void is observed, filled up with crystallized microcline, analogous to that in the groundmass of the rock. In the same void are observed accumulations of small crystalline growths of biotite, as well as small crystals of altered pyrochlore. The specimen of pegmatite consists of a dense aggregate of pinkish-white microcline (partly microcline-perthite with ingrowths of albite), finely foliated accumulations of biotite and areas of crystalline finely-flaky kaolinite (possibly after nepheline). Of rare minerals, grains of brownish and colourless fine zircon, grains of greenish apatite and small grains of brown limonite have been recorded. Among the primary minerals of the pegmatitic body a grain of common reddish-brown, quite unaltered pyrochlore, grown into the microcline, was noted. Thus, the composition of the chief primary minerals of the pegmatitic body and that of the miarolitic cavity are identical.

Not only pyrochlore, but all other minerals of the miarolitic cavity have been affected by secondary alterations. The alterations are manifest in the following*:

1) Microcline-perthite has been corroded and replaced by secondary albite. Albite in the form of white colourless laminated grains corrodes the surface of the

* Precise constants have been determined by the author for all the minerals, both primary and secondary, by the immersion method.

microcline crystals, and covers them with a characteristic imbricate aggregate of flat crystals. The albite is often found to be coated with thin films of brown iron oxides.

2) Biotite is hydrated, converted in small grains in hydrobiotite and coated with films of limonite.

3) The alteration of pyrochlore is very considerable. The corrosion of the crystal faces is distinctly pronounced, the surface of the crystal representing an irregular network of products of decomposition, in the meshes of which corroded areas

of reddish-brown primary pyrochlore are visible (see figure). Small crystals are usually entirely converted into shell-like secondary decomposition products, representing now obvious remains of octahedra with surviving relicts of pyrochlore, and now loose laminated or shapeless aggregates.

The products of decomposition of pyrochlore present an amorphous substance with a lustrous conchoidal fracture, yellow, sometimes slightly greenish in colour, passing to friable powdery cream-coloured varieties. This substance is very friable, easily crumbling into a yellow powder. The composition was studied qualitatively by the microchemical method, since no more than about 4 mg of pure substance could be collected under the binocular. No determination of alkalies could therefore be made. The optical properties were studied by the immersion method.

Optically the products of decomposition of pyrochlore represent a transparent greenish-yellow homogeneous substance, almost colourless in thin splinters. The internal structure is granular, sometimes colloform. The mineral is isotropic. Individual grains show anisotropy, with a very low birefringence. Refractive index very high: $N>2.06$ (about 2.10); in some grains $N=2.06$, that is the refractive index varies within a certain narrow range.

The microchemical study yielded the following results: $(Nb, Ta)_2O_5$ group, very much; SiO_2, $(Ce, La)_2O_3$ group, CaO and H_2O, much; FeO, little; MnO and (Y_2O_3) group, UO_2, ThO_2 and TiO_2, very little and traces; Fe_2O_3, absent.

The data furnished by the optic and qualitative microchemical study of the mineral which is the product of a secondary alteration of pyrochlore, very closely approach those obtained by the present author for the mineral formerly called b.omstrandine in the Ilmen Mts., and identified by the present author as marignacite [1]. The latter, as it has been ascertained, is formed in the Ilmen Mts. also as a result of a replacement of pyrochlore. Marignacite, for the first time described from the pegmatites of Wisconsin (USA)[8], appears to be also a secondary mineral after pyrochlore. In this connexion it is of interest to compare the analysis of typical Ilmen niobpyrochlore with the data obtained for marignacite and the described product of pyrochlore decomposition (Table 1).

Table 1

Oxides	I. Pyrochlore, Ilmen Mts. (after Khrushchev), in %	II. Marignacite, Wisconsin, USA (after Wideman and Lehner), in %	III. Marignacite, Ilmen Mts. mine No. 28 (qualitative microchemical analysis)	IV. Product of alteration of pyrochlore, Ilmen Mts., mine No. 107 (108?) (qualitative microchemical analysis)
Na_2O	3.35	2.52	very little	—
K_2O	0.87	0.57	—	—
MgO	none	0.16	—	—
CaO	14.05	4.10	little	much
MnO	—	none	—	traces
FeO	2.52	0.02	little	little
Fe_2O_3	—	0.50		none
Y_2O_3	0.56	5.07	very little	very little
Er_2O_3		none		
Ce_2O_3	2.16	13.33	much	much
Di_2O_3	1.94	none		
La_2O_3	1.23	»		
UO	2.63	—	little	very little
ThO_2	4.28	0.20	very little	
SiO_2	—	3.10	little	much
TiO	8.32	2.88	»	traces
Nb_2O_5	56.01	55.22	very much	very much
Ta_2O_5	—	5.86		
F	2.77	—	—	—
H_2O	—	6.40	much	much
Total	100.69	99.93		

From the comparison given in Table 1 it will be seen that the results of analyses III and IV are very close to each other, since the products of decomposition of pyrochlore are closely similar to the mineral variety named marignacite. Machatschki[7], who investigated marignacite by the crystal-chemical method and considered it to be a variety of pyrochlore with a replacement of the large atoms of the cation group in the structure of $X_2Z_2(O, OH, F)_7$ by water, has also arrived at conclusions pointing to the secondary nature of this mineral. The author finds it therefore possible to assume a certain variability in the composition of marignacite minerals with regard to the degree of hydration (*i. e.* the ratio of the X group to H_2O), in the ratios of Ca : TR, Si : Ti, in the ratios of U and Th to other cations. The general chemical trend of the conversion of pyrochlore to marignacite manifest in the hydration, in the replacement of Ti by Si (after Machatschki, NbO by SiO_2), the removal of RO (possibly also U and Th) and the enrichment of the molecule with rare earths, is sufficiently well pronounced. There is no room for doubt either regarding the specific genetic relation between pyrochlore and marignacite.

In order to analyse the genetic conditions of formation of marignacite products of alteration of pyrochlore, let us consider the whole complex of alteration of minerals observed by the author in the miarolitic cavity of the hand specimen de-

Table 2

Primary minerals	Secondary minerals	Note
Microcline-perthite	Albite	Supply of Na
Biotite	Hydrobiotite, limonite	Hydration
Pyrochlore	Marignacite products of alteration	Hydration, supply of SiO_2

scribed. On the basi sof the above data, these alterations may be tabulated as above (Table 2).

Thus, when analysing the secondary alterations of all the minerals observed on the miarolitic cavity along with pyrochlore, the alkaline sodium-silica metasomatosis, closely related to hydration, should be considered to be the chief process responsible for the secondary alteration of pyrochlore. This is in full accord with the F geophase of Fersman(3), with its characteristic development of late albitization of feldspars; the genetic position of the conversion of pyrochlore into secondary marignacite alteration products is thus beyond any doubt. The chemical conditions of the processes characterized by fluids (supercritical solutions) of Na, Si ions and the chief volatile mineralizer—the water, should *a priori* have favoured a complete preservation of the $[NbO_3]'$-anion in niobo-titanate, since sodium salts of niobic acid, especially in an alkaline medium, are known to be entirely insoluble. On the contrary, titanium could have been easily leached by hot sodium solutions and, according to the law af acting masses, replaced in the structure of pyrochlore by silicon. The leaching of a part of iron, uranium and calcium cations in this geochemical process cannot be doubted. Thus, the process of «marignacitization» of pyrochlore is associated with the F geophase, with the action of silica-alkaline sodium supercritical solutions, and accompanied by a supply of H_2O, SiO_2, a varying removal of Ti, Ca, Fe, U and Th, and a corresponding enrichment of the residual product—marignacite with the niobium group and rare earths. It seems to the author that from its genetic type this process may be compared with the formation of leucoxenes and the sphenization of ilmenite. In fact, in both cases the alteration products are distinguished by certain variations in the ratio of the principal components (for leucoxene, $CaO : FeO$ and $SiO_2 : TiO_2$) and water(5). On the other hand, the sphenization of ilmenite, as described by Kryzhanovsky and Ginzburg(2), is ascribed by these authors to the development of the pegmatitic process proper and to the corresponding «Nachschub» with a supply or removal of a number of elements, an accumulation of rare earths, *etc.* It is of interest that this process does not affect at all niobium rutile, *i. e.* the $[NbO_3]^{-1}$-ion in the lattice is also stable here.

The existence of secondary alteration of titano-niobates of the group of pyrochlore under the conditions of development of late geophases (E—F) of the pegmatite process has thus been safely established, and its alteration products with certain unavoidable variations in the composition, have been found closely to correspond to the mineral variety described by the name of marignacite.

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REFERENCES

[1] Г. П. Барсанов, Зап. минер. об-ва, LXXIII, No. 2—3 (1944). [2] В. И. Крыжановский и А. И. Гинзбург, Тр. Гос. Ильмен. заповедника, в. II, М. (1940). [3] А. Е. Ферсман, Пегматиты, 1, изд. 3, 1940. [4] К. Д. Хрущев, Зап. Минер. об-ва, сер. 2, 31 (1894). [5] C. Doelter, Handb. d. Mineralchem., III, Abt. 1 (1918). [6] C. Gross, Min Mittel. v. Tscherm. (1881). [7] F. Machatschki, Chemie d. Erde, 7 (1932). [8] S. Wideman and V. Lehner, Amer. J., 23 (1907).

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SOIL SCIENCE

CLASSIFICATION OF IRRIGATED OASES AND THE MOST IMPORTANT MEASURES TO BE TAKEN TO PREVENT SOIL SALINIZATION

By V. A. KOVDA

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According to natural conditions, water and salt regime and balance, conditions of drainage, as well as to the measures to be taken in order to control both their gradual convertion to swamps or salinization, the irrigated oases of the USSR can be subdivided into the following three groups.

1. Oases largely possessing natalural drainage

A. Upper (III—IV) old-alluvial and piedmont plains (mostly upper terraces of the Ferghana valley, the Samarkand and near-Tashkent oases, the Ak-Gasa terrace of the valley of the Vakhsh River, the Syrt plains of the trans-Volga, *etc.*).

The groundwaters are situated at a depth of more than 20—30 m, having a common run-off. Mineralization low or medium.

Balance of groundwaters checked by run-off.

The salt balance tends to desalinization, which process is traceable to a very remote date.

Influence of irrigation. Irrigation contributes to the desalinization of soils, grounds and groundwaters. There is mostly no menace of secondary salinization. Terrains lying along the lower reach of the rivers undergo swamping and salinization because of the vicious practice of water supply.

B. Lower flood-plain and over-flood-plain terraces in the upper and middle reach of rivers (mostly East Ferghana, lower terraces of Chirchik, Angrena, Zeravshan, Surkhan-Daria, *etc.*).

The groundwaters are situated near the surface, frequently wedging-out over considerable areas, though having a common intense run-off corresponding to the slope of the terrain or river. Mineralization poor.

Balance of groundwaters checked by transpiration and mainly by run-off.

Salt balance. Salinization, if any, is limited to accumulation of $CaCo_3$, $CaSO_4$, and of small amounts of Na_4SO_4.

Influence of irrigation. There is neither immediate menace, nor development of salt concentration in the soils. Water losses on filtration or on transgression of the rules of water-supply lead to swamping.

Most important measures to be taken. Control of filtration, rational water-supply and draining of swampy massifs by means of spaced collectors and small drains.

2. Oases largely possessing poor natural drainage

C. Middle (II, III) river terraces and elevated plains (Golodnaya steppe, valleys of the Vakhsh and Arax rivers, second terraces of the trans-Volga regions, *etc.*).

The groundwaters have a poor common run-off and a local run-off conformable to the relief of the depression. Previous to irrigation, depth of groundwater, 5—10 m.

In depression of mesorelief, 2—5 m. Mineralization medium; in depressions of mesorelief, increased.

The balance of groundwaters on the elevated elements of the relief is regulated by way of run-off and transpiration; in undrained depressions, by way of transpiration and evaporation.

Before irrigation the salt balance tends to general desalinization and local salinization of the undrained depressions of the mesorelief; soils and groundwaters are characterized by considerable residual salt contents.

Influence of irrigation. When correctly applied, irrigation contributes to a gradual general desalinization of the land situated at the elevated elements of the relief.

Secondary salinization may develop as a result of a rapid rise of groundwaters in the case of irrational water supply and of a failure to control percolation; it sets in first in soils stretching along channels, and then also at edge-massifs, and especially in undrained depressions. Increased water consumption on the fields and backward agrotechnique are factors contributing to the development of secondary salt concentration.

Most important measures to be taken. Improvement of agrotechnique, extreme thrift in water consumption, control of the feeding of groundwaters at the expense of losses in the irrigation network and on fields. Spaced deep collectors and local draining of undrained solonchak depressions, in case their reclamation be necessary.

D. Lower terraces and plains in the middle and lower reach of rivers (the Volga-Achtuba flood-plain, Shaulder, the Charjow-Farab oase, Central Ferghana, the near-Tugay terrace of the Golodnaya Steppe, *etc.*).

Groundwaters during the major part of the year are headed by the rivers by which they are fed; at the time as the water level is low they are poorly fed by the rivers, approach the surface (1—4 m), and show variable mineralization which may be rather high in depressions.

In regions adjoining the rivers and at elevated spots the balance of groundwaters is regulated by the feeble local run-off and by transpiration; in the central regions and depressions of the relief it is checked by transpiration and evaporation.

Salt balance tends towards a slow general accumulation of salts, while tracts of land lying along river beds and elevated regions become gradually lower in salts.

The central parts of plain-terraces and their depressions undergo intense salt concentration.

Influence of irrigation. Irrigation causes desalinization of the soils of the elevated relief and along the channels, at the expense of local run-off of the soil groundwaters. Under the actual conditions of irrigation technique, the soils become higher in salts at the distal parts of the system, in undrained depressions and vacant land within the oases. In ancient irrigated oases the concentration of salts in the soil attains its maximum in their peripheral parts.

Most important measures to be taken. Rigorous establishment of rational water consumption and advanced agricultural procedures. Digging of spaced deep collectors. In the case the coefficient of useful land (CUL) is increased at the expense of a total reclamation of solonchaks, deep selective drains are to be built. As a means against overflows dikes may be recommended. The severe conditions of the waterlayer balance are somewhat compensated for by the actual low CUL

(0.35—0.50). Any increase in the CUL should be accompanied by adequate measures tending to reduce or divert the groundwaters.

3. Oases, largely undrained

E. Dry deltas (dry deltas of the rivers of Zeravshan: the oases called Bukharsky and Karakulsky, Kashka-Daria, Shirabad-Daria, Sokh, Isphara, Murgab, Tejen, Gheokchay, Bolgarchay, *etc.*).

The groundwaters are fed by rivers, channels; they are desalinized in the upper and central tracts of the dry delta, where they are situated deep and have a run-off. Within the transitional and peripheral zone they are wedging-out in the form of poorly mineralized springs. Within the lower and peripheral parts of dry deltas they nearly outcrop at the surface (1—3 m); they have no run-off and are distinguished by a high degree of mineralization.

The balance of groundwaters in the upper and central parts of the delta, as well as within the zone of wedging, is largely checked by the run-off; in the lower and peripheral parts of the delta it is controlled by transpiration and evaporation, mainly by the latter.

The salt balance of the peripheral parts of the dry delta tends towards progressive salt accumulation; at the upper and central parts of the deltas it is distinguished by a transit of salts, or by a gradual reduction of their concentration.

The lower and peripheral parts of the dry delta are occupied by loamy and clayey meadow-solonchak and clayey swampy-solonchak soils, and by solonchaks.

Within the zone of meadow soils containing no salts or poor in them are prevalent. The upper and central parts of the deltas are occupied by soils of a coarse mechanic constitution, without any trace of salinization.

Influence of irrigation. The effect of irrigation is to increase the process of desalinization of soils and groundwaters of the upper and central parts of dry deltas.

Under the conditions of the irrigation technique actually employed, irrigation goes to increase the accumulation of salts at the peripheral parts of dry deltas, especially if no control of diversion of groundwaters is practised, and if the agrotechnique applied in the region is a rather backward one.

Annual overflows, immoderate water consumption and the violation of the rules of rational water supply, all these are factors likely to lead to the swamping of soils at the periphery of the delta, or to a sharp increase in the concentration of salt in the soils.

The most important measures to be taken. Losses on filtration in the upper and central parts of dry deltas are to be prevented by all possible means. A dense network of irrigation must be developed in the upper parts of the deltas. Within the zone of wedging groundwaters are to be subjected to captation by means of small drains and used for irrigational purposes. At the periphery of the solonchak zone of the delta an emergency-network of collector faults is to be built. At the time being the water salt balance is maintained stable because of a low CUL (0.2—0.3). As soon as the solonchaks of this zone have been converted to orable land, measures to lower or divert the groundwaters will become imperative there.

F. Marine deltas (deltas of the rivers Volga, Terek, Amu-Daria, Syr-Daria, Kura and Arax).

The groundwaters are close to the surface (1.5—2.0 m deep); they are headed and fed the whole year round by the rivers, and in many cases by the sea, too. They have no common run-off and are controlled by local run-off, but mainly by evaporation and transpiration. Their mineralization varies greatly, but is high on the whole.

The salt balance tends towards a general salt accumulation, occasionally accompanied with local migration of salts to vacant land.

The soil mantle is made up of swampy-meadowy and meadowy-solonchak soils, tending to salinization, and of solonchaks.

Influence of irrigation. Massifs irrigated long since and tracts of land along irrigation channels become low in salts at the expense of local run-off and salt migration.

For lack of a common run-off both of groundwaters and salts the natural concentration of salts in soils is intensified, especially in undrained depressions at the peripheral and unirrigated parts of the oases. Intense salt accumulation in irrigated land is increased because of the backward agrotechnique and the lack of crop rotation.

The most important measures to be taken. At the present time the water-salt balance is kept stable by a low CUL (0.3—0.4), which results in the salinization of vacant tracts of land, without entirely preventing the cultivation of land. The increase of the CUL should proceed along with measures tending to decrease and divert the groundwaters (dip collectors, drainage).

Special thriftiness should be prescribed in the consumption of water, as well as a total renouncing to the fault of water in excess, as well as a maximum reduction of losses on filtration. Advanced agricultural procedures tending to eliminate or reduce evaporation are to be imposed. As a means to defend the lowland against annual overflows, the erection of dikes should be prescribed.

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PLANT PHYSIOLOGY

ON CERTAIN ANOMALIES IN THE EMBRYOLOGY OF *POLEMONIUM COERULEUM* L.

By L. G. MARKOVA

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The material for the embryological study of *Polemonium coeruleum* L. was collected from wild growing plants in the vicinity of Tomsk. It was fixed in chrome-acetic-formalin after Navashin's prescription and, partly, with the fixative of Carnoy. The microtome sections were 12—16 μ thick. The material was stained with iron haematoxylin after Heidenhain.

Unicellular archesporium was found in small flower buds, which falls in line with the results by Jonsson (1879/80) and Schnarf (1921). The macrospore mother cell entered normal reduction division, and the formation of macrospore dyads at the first metaphase stage was observed by me.

Jonsson described in *Polemonium coeruleum* the formation of the embryo-sac after the *Scilla*-type, *i. e.* the starting point of development is not a tetrade but a dyad of macrospores; the upper macrospore degenerates, while the lower one, after three divisions, gives rise to an embryo-sac of eight-nucleate type. I could not find in my material all the stages preceding the eight-nucleate embryo-sac, so that the question as to the type of development remains unsolved for the time being.

In the opened flowers the embryo-sac was fully developed and differentiated, the antipodes were three in number, rarely two, and showed ephemeral character, disappearing very quickly. Billings (1901), Jonsson and Schnarf point out the early disappearance of the antipodes in the family *Polemoniaceae*.

In the five-nucleate embryo-sac the egg apparatus was represented by a large egg-cell surrounded with a greatly vacuolized cytoplasm and two stretched synergids which suffered early breakdown. The antipodes were absent. Polar nuclei have already fused together. The embryo-sac attained very considerable size at this stage. In wilting flowers the process of fertilization had evidently already taken place, for in many slides there were visible remnants of the pollen tube. All the anomalies observed by me were confined to this period. The principal point was that in no case division of the egg-cell or developing embryo could be detected, while the fertilized endosperm nucleus entered division, and nuclei of the endosperm of the nuclear type were formed. The same phenomenon was described also by Billings and Schnarf for certain representatives of the family *Polemoniaceae*.

Sometimes the endosperm nuclei were not arranged near the walls but assembled together forming a group of five and more nuclei covered with a common membrane. On other instances the cavity of the embryo-sac was found to hold several large endosperm nuclei containing a great many nucleoli; they also might have originated from the fusion of several nuclei. In one case I observed the formation in the micropylar region of two egg-cells instead of one. In another case a complete

degeneration of the egg-cell was found, while the secondary nucleus was fertilized and ready to enter division.

It might be suggested that when the pollen tube ejaculated its contents into the embryo-sac, no fusion of the one of the sperms with the egg-cell took place. This was corroborated by the fact that I succeeded in observing a sperm lying at the boundary of the egg-cell and not fusing with it, while the large endosperm nucleus was filled with intensely stainable matter and was going to divide.

Such cases were described in certain forms possessing apomictic mode of reproduction, *e. g.* gynogenesis, provided that the egg-cell was able to develop into an embryo without being fertilized.

In many cases in the micropylar part of the embryo-sac there was found an intensely stained large body, containing two black clots. It is possible that it was the second pollen tube which had penetrated the embryo-sac and could not pour out its contents, the black clots being degenerating sperms.

Unfortunately, no later stages were available in the fixed material, which prevents us from reaching a definite conclusion as to the development of the embryo in *Polemonium coeruleum*. However, even from the data reported above one may think that the process of embryogeny is abnormal in this form, parthenocarpy possibly taking place here.

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REFERENCES

[1] Ake Gustafson, Hereditas, XXI, 1 (1935). [2] Fagerlid-Folke, ibid., XXVI, 1—2 (1940). [3] В. Поддубная-Арнольди, Бот. журн. СССР, **25**, No. 1 (1940). [4] K. Schnarf, Handb. der Pflanzenanatomie, **10**, 1929. [5] P. N. Schurhoff, Die Zytologie der Blumenpflanzen, Stuttgart, 1926.

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PHYTOPATHOLOGY

TRANSMISSION OF THE MOSAIC VIRUS OF TOBACCO THROUGH LARVAE OF *PLUSIA GAMMA* L.

By K. S. SUKHOV and A. M. VOVK

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Being highly contagious and resistant to various external influences, the mosaic virus of tobacco might be supposed to be easily transmitted through insects. It has been proved, however, that sucking insects feeding on tobacco do not transmit the mosaic virus. As to the gnawing insects, their rôle in the transmission of the disease has not yet been studied.

We have tested the possibility of transmission of the mosaic virus through larvae of *Plusia gamma*. Seventy larvae were included in the experiment. The larvae had first been put on plants affected with mosaic and then transferred, 3—4 per plant, to healthy tobacco plants. In one set the larvae were left on the plants for several hours; in another experimental set they were removed from the plants immediately after the first injuries had been inflicted to the leaves by their jaws. In this way we intended to prevent the consuming by the larvae of those very regions of the leaves into which the virus was likely to be introduced by them at the initial injury. If the larvae were allowed to feed on the tissues of healthy plants for a longer period of time, their jaws could gradually become free from the virus. It should be noted that many of the larvae, when transferred to the leaves of tobacco, begin by eating the pubescence; this causes a slight traumatization of the plant, similar to that arising when experimental inoculation is practised by rubbing the surface of the leaves with a pestle, or with a piece of fabric moistened with the infecting soap.

Transmission of the Mosaic Virus of Tobacco by *Plusia gamma* Larvae

Type of experiment	Number of experimental plants	Number of plants which have contracted the disease	Percentage of infected plants
Larvae removed immediately after traumatization of healthy plant	23	3	13.0
Larvae left for a long time on healthy plant . . .	15	3	20.0
Total number of experimental plants . . .	38	6	15.70
Total number of control plants	280	3	1.07

As may be seen from the table, the larvae proved to be able to transmit the virus in a considerable number of cases. The proportion of successful inoculations

was somewhat higher when the infected larvae remained on the healthy plant for a long period of time than when they were left on the plants for a short time. In the intestine of the larvae the virus becomes very soon inactivated. Their excrements may, therefore, hardly become a source of infection. The extract from the excrements of the larvae which had for a long time been feeding on diseased tobacco plants, was tested on 8 leaf-halves of *Nicotiana glutinosa*; on the average it caused 86 foci of necroses per one leaf-half. However, 24 hours after the larvae were transferred to healthy plants, no virus could be detected in their excreta. The following experiments have shown that inactivation of the virus actually took place in this case. The larvae were fed for 17 hours on healthy tobacco plants, then kept without food for 7 hours until defecation has ceased altogether; they were subsequently given 750 mg of mosaic leaves which were fully consumed by them within a few hours. The excrements collected after 16 hours were extracted with 4 cm^3 of water and tested on 6 leaf-halves of *N. glutinosa*; on the average, they have caused the formation of 3.73 foci of necroses per one leaf-half. The extract of a control portion of the mosaic leaves has produced on the average 94 necroses per one leaf-half.

Smith has shown ([3]) extracts from larvae of *Protoparce sexta* and some other species to reduce the virulence of the mosaic virus of tobacco. The concentration of the virus in the excrements of the larvae which had been feeding on diseased plants was greatly reduced as compared to that in the leaves. When the extract had been introduced into the haemolymph of a larva, the mosaic virus was found to be completely inactivated.

Plusia gamma belongs to polyphagous insects. About 25 plant species belonging to 23 families are known to serve as food for its larvae. The greatest damage is caused to the crops of flax, hemp and sugar-beet. Many other cultivated plants are also subject to damage; in particular, species of *Solanaceae* susceptible to mosaic of tobacco: henbane, belladonna, tobacco and red pepper. One must not think that the degree of transmission of the mosaic virus by *P. gamma* is extensive on tobacco plantations. The spread of the virus is undoubtedly increased by many agrotechnical operations, such as picking, removal of side shoots, pulling leaves, *etc.* One may think, however, that there probably exist a number of other transmitters of the mosaic, for example, *Euxoa segetum* Schiff., *Agrotis obesa* B., and *Heliothis armigera* Hb. The first species is a mass pest of the tobacco crop in the Crimea ([1]).

The finding of an insect that is able to transmit the tobacco mosaic is of prime importance, sheding light upon the possible natural mode of transmission of such viruses as the cucumber virus No. 2 and the X-virus of potato, for which no insect transmitters are known up to the present time.

In concluding we express our gratitude to Professor Fedotov for the identification of the larvae and moths which were the object of our study.

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REFERENCES

[1] Ю. Ю. Скалов, Вредители табака в Крымской АССР, Краснодар, 1939. [2] J. A. Hoggan, J. Agr. Res., 49, No. 12 (1934). [3] K. M. Smith, Parasitology, 33, No. 1 (1941); Rev. Appl. Myc., 20, No. 6, 270 (1941).

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EXPERIMENTAL MORPHOLOGY

SELF-DIFFERENTIATION OF SUCKER FROM MATERIAL OF CHORDAL RUDIMENT

By T. DETLAFF

(Communicated by I. I. Schmalhausen, Member of the Academy, 25.IV.1945)

The development of the sucker proceeds in a rather peculiar way: it is self-differentiated from different isolated parts of the ectoderm, without any influence of the other parts of the embryo, either *in vitro* ([1],[2]) or in *vivo* ([3],[4]), and as in non-gastrulating united abdominal halves of gastrulae ([5]) as well in the ectodermal part of exogastrula ([6]). Thus, the sucker arises by means of self-differentiation from ectodermal material which, when developing typically, never gives rise to the development of a sucker. In the experiments described below we observed self-differentiation of the sucker from the external layer of the rudiment of the presumptive chorda. The development of the sucker was not the subject of the present study, the facts here reported being observed in the course of another work.

The experiments consisted in separating the ectodermal layers within the presumptive rudiment of the chorda at the early gastrula stage in *Rana temporaria*, *Bufo vulgaris* and *Bufo viridis*, and in a homoplastic transplantation of the external and corresponding internal layers into the throat cavity of two separate embryos, at the stage of early caudal bud. In the second experimental set the same embryonic material was transplanted into the blastocoel at the early gastrula stage. When the external layer of the chordal rudiment was transplanted into the throat cavity, alongside with the development of irregular epidermis, the differentiation of groups of glandular sucker cells was observed in 57 per cent of cases (11 cases out of 19). In some cases whole transplants were transformed into suckers, the latter being in two cases nearly typical in their structure. In none of the cases (out of 24) corresponding internal layer transplanted gave rise to the development either of sucker or epithelium, but only chorda and in some cases, nervous tissue.

Similar relations were observed when identical material was transplanted into the blastocoel, except that here the relations were not so demonstrative, because the transplantations were homoplastic ones. Sucker structures, or a whole sucker developed in 43 per cent of cases (in 13 cases out of 30). In a number of cases the transplant was found to lie on the surface. It was not harmoniously included into the embryo covering, and gave rise to a sucker structure. Four embryos were found to possess a whole sucker not typically developed and located on the abdominal or lateral surface of the embryo. In another embryo, separate groups of sucker cells were present.

In a number of cases, where the external layer of the chordal region was from the very first superficially transplanted in place of the abdominal epithelium of the gastrula previously removed, the material transplanted was smoothly united to the underlying tissues and formed the epithelium of the covering. A similar

behaviour of the whole ectodermal layer transplanted was observed by Jamada. In a number of cases there was also present a small chorda, or single chordal cells. These experiments showed that the sucker differentiation may occur not only in the throat cavity, and is, therefore, not determined by the special influence coming from this region.

In none of the cases of this experimental set did the internal layer of the chordal rudiment transplanted into the blastocoel lead to a development either of epithelial structures, or of glandular cells of the sucker. In 4 cases only out of 37 the induction of an extra-sucker as a part of a secondary embryo induced from the host ectoderm took place. *In vitro* development of the sucker from a whole chordal rudiment in *Anura* was described by Holtfreter ([2]). His explanation of this phenomenon was that the ectoderm arisen by means of regulation underwent inductive influence from the remaining part of the rudiment. Holtfreter has also pointed out that when the external layer of the chordal rudiment was isolated, it led in some cases to the formation of epidermis alone.

On the basis of the facts described above, we may say that the development of the sucker from the external layer of presumptive chordal rudiment may proceed by way of self-differentiation. In an isolated ectoderm it is apparently self-differentiated from the external layer also, since both in the typical development ([7]) and in xenoplastic transplantations ([8]) the sucker was observed to develop from the external layer of the ectoderm. On the basis of our observations we are inclined to accept Holftreter's point of view and to suggest that the whole ectoderm including the epithelial part of chordomesodermal rudiment, when isolated from the mesoderm, is capable to differentiate into a sucker. We must not be surprised at the striking fact of self-differentiation of the sucker from the external layer of relatively determined chordal rudiment, since the external layer of the chordal part in *Anura* does apparently form the hypochordal plate rather than the chorda. In addition, the peculiarities of the layer in question make it more like the ectoderm than the chordomesoderm.

At definite stages of development the whole ectoderm is capable to develop a nervous system, a lens, an optic vesicle, supporting threads or suckers. However, the development of the sucker experimentally induced takes place without the influence coming from any other part of the embryo. It is not possible to exclude without special investigation that poorely specific influences of external environments may play an inductive rôle in this case. Holftreter's observations ([9]) on the development of a superficial protoplasmic membrane, as well as the inhibiting influence of amniotic fluid upon chitinization of insect integument ([10]) show that the differentiation of the ectoderm may be directly conditioned by the external environments. On the other hand, it is no less probable that the ability to form glandular sucker structures, being in itself a hereditary property, is realized in a similar tissue environment (in the external layer of ectoderm and chordomesodermal rudiment), and does not need any additional specific influences for its manifestation. Holftreter is right in believing that in the case of normal development this ectodermal tendency is inhibited apparently by the influence of the mesodermal underlayer. This suggestion is proved by our experimental data. In two cases the external layer of the chordal rudiment including the mesenchyme cells was transplanted into the throat cavity. The transplants formed epithelial vesicles having normal one-layered epithelium with a basal membrane and a friable mesenchyme inside. The differentiation of the sucker is also inhibited by the chorda present.

No sucker development was observed in the experiments in which whole chordal rudiments (without dividing it into layers) were transplanted into the throat cavity. This inhibitory effect of the mesoderm explains why the sucker fails to form in case the ectoderm is transplanted on to the head or the abdomen([11]), and does not arise there in the course of normal development. It has already been shown, however, that a sucker may develop in this case also, provided that the external layer transplanted be poorly united with the underlying tissue, being isolated directly or by another tissue, *viz.* the epithelium.

As to the region of mandibular arch where the sucker is normally developed, the conditions that arise here probably serve to isolate the ectoderm from the underlying mesodermal tissue. To my opinion, this rôle is played by the entoderm, which, as shown by Holftreter in his explantation experiments, unlike the mesoderm does not inhibit the sucker differentiation of ectoderm but rather intensifies it. Thus, the localizing rôle should be attributed to the entoderm of the mandibular region. This suggestion seems to be still more probable in the light of Jamada's investigations ([12]), who showed that at these stages of development the head gut underlies the region of the future sucker, and that the reduction of the former leads to a corresponding decreasing of the latter. Jamada himself supposed that the inductive influence of the head gut was responsible for the results observed by him. However, it would be more correct to admit that the entoderm is responsible for the isolation against the inhibitory influence of the mesoderm. As to the problem of special inductive action of the mandibular region ([9],[11],[12]), it must be said that we have so far no reason either to deny or to accept it, since the whole ectoderm is capable of sucker formation without the influence in question. The typical formation of a sucker at an appropriate place and usually an atypical one when isolated ([2]) cannot be considered as a proof of the specific local influence present here, first, because cases are known where the typical development of the sucker took place under conditions of isolation (exogastrula), and, secondly, because the factors conditioning the formation of a typical organ may vary greatly, often involving a whole complex of processes typical of a given place of the embryo and of a given stage of its development.

The similarity of tissue relations in the mandibular region of *Anura* and *Urodela* may explain the development of sucker in xenoplastic transplantations when *Anura* ectoderm is transplanted into the region of the mandibular arch of *Urodela*. It is needless to assume the presence of a common inductor in the development of the sucker and of supporting threads ([9], [11]).

Consequently, the idea as to the homology of the latter loses its only basis. True enough, Schotte and Edds ([11]) observed the development of a sucker at the place of supporting threads. However, their experiments would be conclusive if only they had obtained the development of supporting threads from the ectoderm of *Urodela* transplanted into the sucker region of *Anura*. But no supporting threads are formed here ([12]), and the entoderm does even suppress the development of supporting threads in the central part of mandibular region ([14]). Thus, entoderm is favouring the differentiation of suckers ([2]) and inhibiting the development of supporting threads ([14]), while the mesoderm inhibits the development of suckers ([2]), being evidently necessary for the development of supporting threads ([14]).

These facts point to a difference rather than to a similarity. The suckers and supporting threads are analogous organs, but there are no morphological or phylogenetic grounds to regard them as homologous ones ([8], [13]). According to Lieberkind ([7]), the suckers have arisen in different *Anamnia* independently of one another, from the ectoderm or entoderm, respectively. Therefore, one may speak only about their convergent development, but not their homology.

The development of suckers is a peculiar case of organogenesis where the hereditary factors of the material are sufficient for the development of a specific glandular differentiation of the sucker. Under definite tissue surroundings (the external ectodermal layer or chordomesoderm) the hereditary factors are able to produce self-differentiation of sucker cells. Under other tissue surroundings these factors are evidently unable to manifest themselves, and neither the mesoderm nor the chordal rudiment proper give rise to a sucker. Thus, the determination of the sucker differentiation, taking place at an early gastrula stage, is realized, like that of the entoderm and chordomesodermal rudiment, in a whole cell layer on the basis of gradient differences of the embryonic layers, and probably also on the basis of their definitive interrelation with the external surroundings. Holtfreter has pointed out that in the case of typical development the surrounding parts of an embryo inhibit the differentiation of a sucker everywhere, except the place of its normal development. Being in contact with the entoderm, the

ectoderm becomes isolated from the inhibitory effect of the mesoderm. In other words, the interaction between the parts of an embryo plays a clearly expressed localizing rôle, and is mainly conditioned by an inhibitory action rather than by a stimulating one, which takes place in the case of inductors. The possible significance of the inhibitory influences during ontogenesis was pointed out by Holtfreter, Balinsky([14]), and, especially, by Schmalhausen ([16]). According to Schmalhausen's point of view, developed in his system of correlative interrelations, the interaction of the inductive type is no more than a particular case. The development of a sucker is another evidence in favour of this point of view.

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REFERENCES

[1] J. Holtfreter, Roux'Arch., **124** (1931). [2] Idem, Roux'Arch., **138**, No. 4 (1938). [3] Bytinski-Salz, Roux' Arch., **114** (1929). [4] H. S. Emerson, J. Exp. Zool., **90**, No. 3 (1942). [5] Д. П. Филатов, Биол. журн., VI, № 2 (1937). [6] J. Holtfreter, Roux'Arch., **129** (1933). [7] Ingvald Lieberkind, Vergleichende Studien über die Morphologie und Histogenese der larvalen Haftorgane bei den Amphibien, 1937. [8] H. Spemann u. O. E. Schotte, Naturwiss., **20**, 463 (1935). [9] J. Holtfreter, J. Exp. Zool., **93**, No. 2 (1943). [10] П. П. Иванов, Общая и сравнительная эмбриология, 1937. [11] O. E. Schotte and Mac. V. Edds, J. Exp. Zool., **84**, No. 2 (1930). [12] T. Jamada, J. Fac. Sci. Imp. Univ. Tokyo, **5**. 1 (1938). [13] E. Rotmann, Verh. d. Zool. Ges., S. 76 (1935). [14] B. I. Balinski, Acad. Sci. of the Ukrainian SSR, No. 12 (1938). [15] H. Spemann, Experimentelle Beiträge zu einer Theorie der Entwicklung, 1936. [16] И. И. Шмальгаузен, Природа, 1945. [17] T. Jamada, J. Fac. Sci. Imp. Univ. Tokyo, IV, No. 3 (1933).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
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PHYSIOLOGIE

SUR LE RÔLE DES SUBSTANCES ERYTHRODIÉRETIQUES DANS LE PHÉNOMÈNE DE LA RÉGÉNÉRATION DU SANG

Par J. G. OUJANSKY

(Présenté par L. A. Orbeli, de l'Académie, le 19. V. 1945)

Nous avons trouvé en 1932(1) qu'une saignée pratiquée à l'un de deux rats en parabiose provoque chez son partenaire l'apparition de l'anémie, anémie présentant même une intensité plus grande que chez l'animal saigné. Ce phénomène pouvait être expliqué par le simple passage du sang de l'animal intact à l'animal saigné: il est en effet démontré que les hématies peuvent réellement traverser une anastomose capillaire chez des animaux en parabiose (2). Ceci semblait d'autant plus vraisemblable que, dans les conditions de dérèglement aigu de la circulation sanguine par suite de la saignée, ce passage pouvait prendre une importance particulière.

Il se montra cependant par la suite que le phénomène étudié était considérablement plus complexe qu'il ne l'avait semblé de prime-abord. Nous avons montré dans une série de recherches (3—6), qu'à la régénération intensifiée des erythrocytes précède une destruction accrue de ces derniers. Il se révéla en outre que le serum de l'animal saigné est susceptible, suivant la quantité injectée et les particularités de l'animal récepteur, de provoquer chez celui-ci aussi bien un phénomène d'hyper- que d'hypoglobulie (7—8).

Ainsi donc, le phénomène que nous avons décrit, pouvait également s'expliquer par une influence humorale de l'animal saigné sur le tissu hémopoïétique ou les hématies de son partenaire.

Cette explication devint particulièrement plausible après les travaux de Kolpakov (10), qui put exclure des conditions de génèse du phénomène décrit, l'influence due au dérèglemant aigu de la circulation sanguine par suite de la saignée.

S'écartant en celà de notre méthode, il saignait l'animal d'expérience avant de l'unir par anastomose à son partenaire normal, mais même en procédant de cette manière il obtint des résultats tout à fait analogues aux notres.

La présente étude se propose d'éclaircir la nature de cette influence humorale de l'animal saigné sur son partenaire normal; la question se posait de savoir, si l'anémie chez le dernier n'était pas due à une destruction intensifiée de ses hématies, et au cas où ceci devait se confirmer, de déterminer les relations unissant cette hémolyse au phénomène de la renovation du sang chez l'animal saigné.

Les expériences portèrent sur des rats blancs en parabiose, unis entre eux par anastomose cutanéo-musculaire (11). Une biopsie de la rate avant la saignée permettait un contrôle histologique; on déterminait en outre le nombre d'erythrocytes et la quantité d'hémoglobine. Ce n'est qu'ensuite que l'un des animaux anastomosés était soumis à une saignée des vaisseaux de la queue. Cette saignée était de l'ordre du $^1/_3$ de la quantité totale de sang de l'animal (en admettant que le poids

du sang représente le vingtième du poids du corps). Une nouvelle analyse de sang était faite dans la période post-hémorragique. Les animaux étaient ensuite sacrifiés au bout d'un temps plus ou moins long après la saignée. Les rates étaient soumises à une étude histologique du point de vue de l'erythrophagocytose, de l'erythrodiérèse et du dépôt de pigment ferreux*.

Toutes ces expériences, de même que nos recherches antérieures sur les hémorragies provoquées chez des rats en parabiose [1], ainsi que les expériences de Kolpakov[10], ont montré que l'animal intact présentait une anémie même plus intense que son partenaire en parabiose soumis à la saignée.

La destruction des hématies présente également un caractère de régularité dans ses différentes phases. Si les animaux d'expérience étaient sacrifiés dans les deux premiers jours suivant la saignée, la destruction des hématies chez le partenaire se manifestait par une intensification de l'erythrophagocytose. Si, par contre, on laissait passer quatre jours et plus entre la saignée et la mise à mort des animaux, un dépôt intensifié de pigment ferreux remplaçait l'activation de l'erythrophagocytose. Chez l'animal saigné la destruction post-hémorragique des erythrocytes se passait exactement de la même manière que chez des animaux non-anastomosés, ainsi que nous avons pu l'observer [3,4], mais le phénomène était moins marqué chez lui que chez son partenaire non-soumis à la saignée. Ainsi donc les variations du nombre d' erythrocytes chez des animaux en parabiose reflètent le processus de destruction de leur hématies.

Nous avons obtenu des résultats analogues dans nos expériences de parabiose, en unissant par anastomose des animaux préalablement saignés à des animaux intacts.

Dans ces cas ces derniers présentaient de l'anémie dès les premiers jours de leur mise en parabiose avec les animaux saignés.

Supposant d'après les faits rassemblés que l'intensification dans la destruction des hématies au cours de la renovation du sang est génétiquement en rapport avec l'activation de l'erythropoïèse, nous avons procédé à la série d'expériences suivantes. Des rats normaux étaient mis en parabiose avec des animaux à hémopoïèse, activée par maintien prolongé de ces derniers dans une atmosphère à pression réduite. De cette manière on provoquait chez eux une augmentation de 20 à 30 pour cent du nombre de globules rouges et du taux d'hémoglobine.

Il s'est montré que l'animal à hémopoïèse intensifiée provoque également chez son partenaire normal un phénomène d'hypoglobulie, rappelant ce qui avait lieu dans les expériences de parabiose avec un partenaire soumis à la saignée. Rapprochant ce phénomène de nos expériences précédentes, on peut admettre par analogie que dans ce cas également, l'hypoglobulie est due à une intensification dans la destruction des hématies.

A notre avis, les faits exposés permettent d'élargir sensiblement nos conceptions du méchanisme de la rénovation sanguine, et donnent en particulier une explication nouvelle de l'activité hémopoïétique du sérum des animaux saignés [sérum de Carnot [12,13]] ou des animaux maintenus dans une atmosphère à pression réduite[9].

L'injection à l'animal d'un tel sérum a vraisemblablement pour effet de provoquer toujours au début une intensification du processus de destruction des hématies, et ce n'est qu'ensuite, secondairement, que se produit une stimulation de l'erythropoïèse. Mais dans les expériences caractérisées par une injection unique d'un tel sérum (expériences de Carnot et Deflendre[12] et de beaucoup d'autres) l'effet hypoglobulique est souvent masqué par la régénération sanguine consécutive.

Par contre, dans les expériences de parabiose, quand on soumet l'animal à l'influence continue et prolongée de substances erythrodiérétiques, le processus de destruction des hématies se manifeste avec une grande intensité.

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* Comme l'avaient démontré nos expériences antérieures, la destruction des erythrocytes chez les rats a lieu principalement dans la rate.

Il semble que la plupart des facteurs stimulant l'erythropoïèse exercent leur action indirectement par l'intermédiaire d'une phase d'activation de la destruction des hématies. Et ce sont vraisemblablement les produits de destruction cellulaire des globules rouges qui constituent principalement les facteurs stimulant directement l'erythropoïèse. Une telle manière de voir est corroborée par le fait que le «manque d'oxygène», ce facteur bien-connu stimulant l'erythropoïèse, provoque également au début de son action une destruction accrue des hématies.

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LITTÉRATURE CITÉE

[1] Я. Г. Ужанский, Журн. медичн. цик. У. А. Н., **2**, 1, 57 (1932). [2] Я. С. Андросов и Н. В. Ромодановская, Труды 2-го Ленингр. мед. ин-та, вып. VI, 13 (1935). [3] Я. Г. Ужанский, Журн. медич. цик. У. А. Н., **2**, 2, 313 (1932). [4] Он же, там же, **2**, 4, 663 (1932). [5] Он же, Врачебное дело, 23—24 (1932). [6] Он же. Архив пат., анат. и пат. физиол., VI, 4, 88 (1940). [7] A. Zih, Endokrinologie, **3**, 81 (1929). [8] A. Zih, Pflügers Archiv, **225**, 613 (1930). [9] Я. Г. Ужанский, Мед. журн. У. А. Н., V, 3 (1935). [10] И. В. Колпаков, Юбил. сборн. труд. Астрах. мед. ин-та, 1938. [11] Л. Р. Перельман, И. В. Иконен и Н. В. Ромодановская, Труды Ленингр. туберк. ин-та, Юбил. сборник, 1933. [12] P. Carnot et Cl. Deflendre, C. R., **143**, 384 (1906). [13] P. Carnot, C. R. Soc. Biol., **58**, 344 (1906).

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HELMINTHOLOGY

A CONTRIBUTION TO THE REARRANGEMENT IN THE TAXONOMY OF TREMATODES BELONGING TO THE *OPISTHORCHIDAE* FAMILY

By K. I. SKRJABIN, Member of the Academy

I. On the Subdivision of *Diasia* Travassos, 1922, into Independent Generic Components

Travassos (1) had founded the genus *Diasia* for the trematodes *Diasia diasi* of the family *Opisthorchidae*, parasitic of the pancreas of the bird *Plotus anhinga*, which he rightly referred to the subfamily *Pachytrematinae*. Ejsmont (2) committed a mistake by referring to the genus *Diasia* a trematode which had been described by Linton (3) from the pancreas of the bird *Gavia immer* by the name of *Haematotrephus fodiens*. The present author has newly established (1945) that this trematode should, as a matter of fact, be separated into an independent genus for which he has suggested the name of *Erschoviorchis*, and which is a typical representative of the subfamily *Opisthorchinae*.

Olsen (5) has described by the name of *Diasia podilymbae* a trematode from the mesentery of the bird *Podilymbus podiceps*.

While analysing the position of this parasite in the trematode system, I was brought to the following conclusions.

1. *Diasia podilymbae* Olsen, 1938, has no bearing at all upon the genus *Diasia* Travassos, 1922. This is evidenced by: a) the elongated-stretched shape of its body showing parallel lateral edges and an obtuse caudal end; b) the broad caliber of the intestinal trunks and a different type of the oesophagus: c) the median arrangement of the testes which are, besides, situated one behind another; d) the diffuse nature of the yolk-bodies, instead of the star-like arrangement, and a topography of their own; e) the structure of the uterus, and a number of other less important characters. On account of what has been said above, *Diasia podilymbae* should be excluded from the genus *Diasia*.

2. The structure of *D. podilymbae* is so peculiar and so unlike that of any other trematode that this trematode should duly be separated out as a special genus which I give the name of *Plotnikovia* nov. gen., in honour of the Russian helminthologist N. N. Plotnikov whose achievements in the development of the problem involved in opisthorchosis are of great importance.

3. The genus *Plotnikovia* cannot possibly be included into any of the subfamilies founded within the family *Opisthorchidae*. It is a representative of an independent new subfamily, *Plotnikovinae* nov. subfam.

II. To the Taxonomic Position of the Genus *Witenbergia* Vaz, 1932

Vaz (4) has described by the name of *Witenbergia witenbergi* a trematode, parasitic of the small intestine of the freshwater fish *Pseudoplatistoma tigrinum*, which he classed among the subfamily *Opisthorchinae*.

After having considered the position of this trematode in the trematode system in the light of the up-to-date taxonomic conceptions, I feel obliged to pronounce the judgment of Vaz as lacking any solid foundation. *Witenbergia witenbergi* should by no means be referred to the subfamily *Opisthorchinae*, for it differs from all the numerous representatives of the family *Opisthorchidae* by two fundamental peculiarities, *viz.*: 1) an exceedingly elongated prepharynx whose length makes up $^1/_4$—$^1/_5$ of the total body length; owing to this, the pharynx is shifted to the level of the boundary between the first and the second fourths of body length, and 2) the position of the spermatheca which is situated anteriorly of the ovary. Besides, owing to its adaptation to the life in the host's intestine, this genus is armed with a powerful fixatory apparatus, in particular, with an oral sucker equipped with a well-developed musculature, and with a deep body cavity, which enables this parasite to catch and pinch a considerable area of the mucous of the host's intestine.

While I do fully agree with Vaz as to the generic independence of *Witenbergia* I at the same time believe that it was an error on the part of this author to include the above-named genus into the subfamily *Orchipedinae*. The peculiar features in the structure of *Witenbergia* undoubtedly pass beyond the scope of the generic characters, and should, therefore, be raised to the significance of subfamilial taxonomic coefficients.

The complex of taxonomic characters of *Witenbergia* seems to me to display a close affinity with the genus *Phocitrema* Goto et Ozaki, 1930, parasitic of the intestine of the seal. The two genera show an analogous structure of their digestive organs and similar topographic relations between their ovaries and spermathecae. Hence, it would seem but natural to include the genus *Witenbergia* in the same subfamily with *Phocitrema*, *viz. Phocitrematinae*.

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13. VIII. 1945.

REFERENCES

[1] L. Travassos, Folha Medica, Brasil, **3**, 187 (1922). [2] M. Ejsmont, Acad. Polonaise Sc., Lett., Cracovie, No. 6, 7 (1931). [3] E. Linton, Proc. U. S. Nat. Mus., **73**, 1 (1928). [4] Z. Vas, Trabalho do Inst. Biologico Sao Paulo, p. 38 (1932). [5] O. Olsen, J. of Parasit., **24**, No. 3, 215 (1938).